A STEEL PICTURE-TUBE FOR TELEVISION RECEPTION

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with an introduction by J. G. W. MULDER.   621.385.832:621.397.62

A milestone in the development of larger picture-tubes for direct viewing was reached with the tube which appeared on the market some years ago in America, and recently also in Europe, namely the tube with a conical portion made of metal. One of the many advantages of this is that the glass window may be made thinner, thereby permitting the use of drawn glass; this latter has a uniform thickness, which results in an improvement in image quality.

The first types of this new form of tube had a cone made from chrome-iron, a rather expensive material. It may therefore be considered a great step forward that success has been achieved in producing a type of glass which can be sealed to steel, and that at the same time a good solution has been found for the difficult corrosion problem involved.

A tube with a steel cone and a window made of the new glass is now coming into production. The authors describe the development of this tube. Their description is preceded by an introduction dealing with the history of metal discharge tubes.

INTRODUCTION

It has long been a cherished desire to make the walls of various discharge tubes as much as possible from metal and to use glass only where necessary for the insulation of the leads. New reasons for such a development are constantly arising: in transmitting valves it is attractive to allow the metal wall to serve as the anode, which can then be intensively cooled by water or by air; in the case of X-ray tubes operating in air, a metal wall is not only attractive for cooling purposes, but gives in addition good protection from undesired radiation and, if earthed, from the high-voltage danger; in industrial rectifiers metal tubes are more suitable than large more vulnerable glass tubes, etc. Similarly, in the case of picture-tubes for television reception — at least of the larger tubes for direct viewing — a metal wall has particular advantages, as will be seen in the following.

In order to realise this objective one must in the first place have available a metal and a glass which can be sealed together to give a vacuum-tight joint; in addition the seal must be able to withstand the temperature changes and localised temperature differences which may arise during the operation of the tube. Less fundamental, yet in practice important, further requirements are that the metal shall not be too difficult to work, not too expensive, etc.

Acceptable solutions have been found in many cases. By the Housekeeper method, for example, in spite of considerable difference in expansion coefficient between glass and metal, a tight seal is obtained by the use of a ductile metal, such as copper, which is made very thin at the place of sealing. The metal can then, by yielding slightly, accommodate its dimensions to those of the glass. Although this seal forms a weak spot, from a mechanical viewpoint, it is nevertheless employed in some types of transmitting valves.

Mechanically stronger is the seal between chrome-iron and lead-glass — two materials which match each other very well as far as expan-

sion coefficient is concerned [1]. Use is made of chrome-iron in numerous tubes manufactured by Philips, either for the contact-pins or for the tube wall itself. This latter is the case in various transmitting valves, in X-ray tubes operating in air, and in some rectifiers and thyratrons (see the final reference quoted in note [1]).

There are certain objections to the use of chrome-iron for the wall, namely the poor workability and the costliness of this material. The same objections also obtain in the case of alloys such as fernico, which has a lower expansion coefficient, matching the harder glass types. In view of the constant demand for metal tubes, efforts have been made to overcome these disadvantages, by attempting to use as the metal a cheap and easily workable type of steel. Two lines of approach have been followed by Philips with this end in view. One approach has led to a particular seal construction, the other to a new type of glass, matching steel in expansion coefficient.

Compression seal

The first solution arose out of the fact that, in general, glass withstands compression better than tension (the ratio of the maximum permissible compression stress to the maximum permissible tensile stress can be as much as 20 : 1) and that it is in fact the tensile strain, with unsuitable construction methods, which is normally responsible for cracks. Because of this conviction we developed in 1940-41 [2]) a rectifier tube (fig. 1a), in which it was arranged that under normal circumstances the glass was under compression ("compression seal"). A seamless steel tube is shrunk around two lead-glass pressed forms (fig. 1b), and serves as the tube wall [3]). At room-temperature the glass is under considerable compression. If the tube heats up during operation, it is true, this compression decreases, as the expansion coefficient of steel (13.0×10^{-6} per °C) is greater than that of lead-glass (9.5×10^{-6} per °C). Nevertheless there is a considerable useful range before the temperature reaches such a value that the compression strain changes over into one of tension, and the chance of cracking becomes significant.
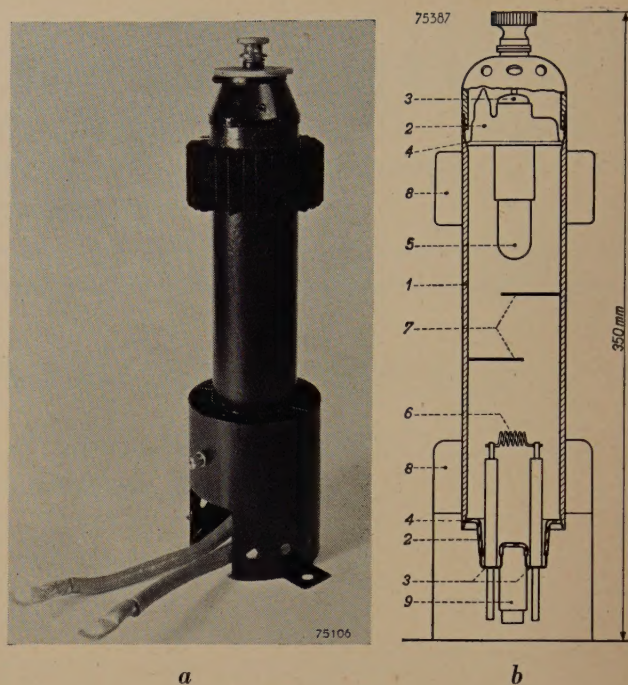

a   *b*

Fig. 1. Steel-walled rectifier tube with compression seals. a) photograph, b) cross-section of the valve made in 1940-1941. The tube wall consists of seamless steel tubing 1; the ends 2 are of pressed lead-glass, with chrome iron caps 3 for the current leads. At the seal 4 between steel and glass, the glass is under compression. 5 anode, 6 directly heated oxide-coated cathode, 7 partitions, which locally constrict the current path and thereby hinder arcing-back, 8 cooling fins, 9 mercury reservoir.
The tube contains mercury vapour and argon, and under a tension of some hundreds of volts is capable of delivering a a current of 15 A average value with a peak value of 75 A.

Since the glass flows, more or less, it relieves itself of strain to a certain extent in time. Over a number of years we have followed the course of the compression strains in a series of samples of this compression seal. The results (*fig. 2*) show that with our construction the maintenance of compression strain is assured for many years.
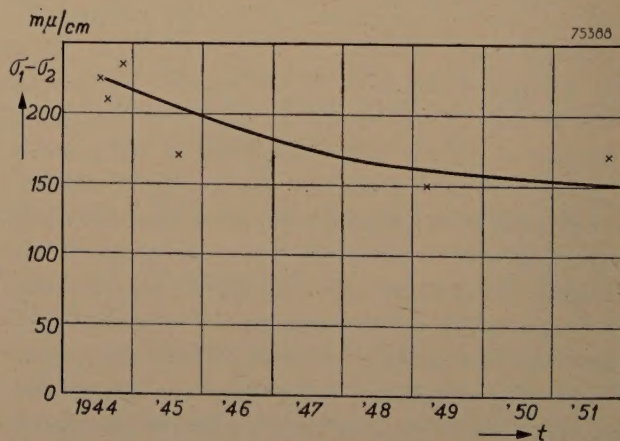

Fig. 2. Difference between two principal strains (tangential and radial), $\sigma_1 - \sigma_2$, in the glass of a compression seal, as a function of the time t over some years. $\sigma_1 - \sigma_2$ is expressed in mμ/cm, the unit which is customarily employed for strain measurements in glass (see note [11]).

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[1]) B. van der Pol, Metal transmitting triodes for high power, Sterkstroom I, 265-266, 1923 (in Dutch); A. Bouwers, A new X-ray tube, Physica 4, 173-179, 1924 and A New Metal X-ray rube, Fortschr. Röntgenstr. 33, 575, 1925. A summary is given in an article by H. J. Meerkamp van Embden, Joints between metal and glass, Philips tech. Rev. 2, 306-312 1937.

[2]) For reasons outside the scope of this discussion, this tube was never brought on the market.

[3]) A procedure for making a compression seal between a steel pipe and pressed glass is described in Dutch patent application No. 102 638.

"Iron" glass

The other avenue followed was the search for a type of glass with approximately the same expansion coefficient as mild steel. The Philips Glass Factory in Eindhoven paid considerable attention to this project. This led in 1938 to a new type of glass (followed later by still more), which conformed to the requirement of sealability to iron and was

may arise at places where hot and cold parts of the tube border on each other. For this reason there is still a preference in the case of rectifiers and other heavily loaded valves for the combination chrome-iron/lead-glass (expansion coefficient $a = 10\cdot5 \times 10^{-6}$ and $9\cdot5 \times 10^{-6}$ per °C respectively) or still better, for fernico/hard glass ($a = 4\cdot6 \times 10^{-6}$ per °C), rather than to the combination steel/"iron" glass
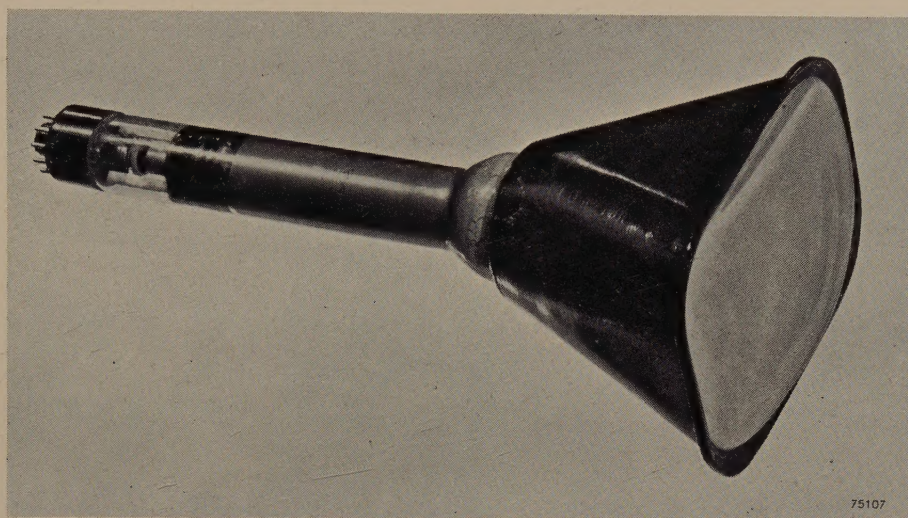

Fig. 3. Prototype picture tube made in 1949. Cone of mild steel, remaining parts of the bulb of (lead-containing) "iron" glass. Diagonal of the window approximately 15 cm.

therefore called "iron" glass. Rectifier tubes were constructed from this glass and mild steel in 1939.

Rectifiers, however, during operation or during warming-up and cooling-down, are subjected to large localised temperature differences. In such cases it is not sufficient that the glass and metal should have practically the same expansion coefficient, but the coefficient must in itself be fairly small; otherwise prohibitively large stresses

($a = 13\cdot0 \times 10^{-6}$ and $11\cdot6 \times 10^{-6}$ per °C respectively).

Conditions are different with picture-tubes, since in this case no large temperature differences arise. *Fig. 3* is a photograph of a picture-tube manufactured in 1949, with a cone of mild steel and a window of "iron" glass.

The following article describes how the development thus introduced was successfully completed with the marketing of a modern steel picture-tube with a diagonal of 43 cm.

Picture-tubes for direct viewing

The first television receivers produced a picture of small dimensions according to present conceptions. In the years 1936-39 the standard tube in Europe had a diameter of about 22 cm; only in "de luxe" receivers was a tube of 31 cm employed.

After 1945 there arose a strong demand for larger images. These are less tiring for the eyes, and at the same time enable a larger company to view the picture. A change was made simultaneously to tubes with a more or less rectangular window

(fig. 3); in the case of large tubes, particularly, this permits a considerable saving in the dimensions of the cabinet.

There are two methods by which the demand for larger images may be met: the construction of larger tubes, and the use of an optical system which projects a considerable enlargement of the image produced on a small tube onto a screen (projection-viewing). Both methods are applied by Philips, but in this article we limit ourselves to the former ("direct view").

With increasing dimensions of (glass) picture-tubes, a number of objections become more and more serious:

1) In order to make the bulb able to withstand the pressure of the atmosphere, the tube window must either be made very thick or be given a strong curvature. In both cases the picture is deformed, so that the maximum usable picture dimensions do not increase proportionally to the diameter (or the diagonal) of the window. Even the

cone of chrome iron. In the course of this year, however, manufacture will go over to an externally similar tube with steel cone. This latter tube is the special subject of this article and is shown in *fig. 4*.

Special properties of metal picture-tubes

As we shall see later, a metal cone permits the use of a window which is fairly thin (5.5 mm) and which needs to be only slightly curved (radius of curvature approx. 75 cm) in spite of the force of
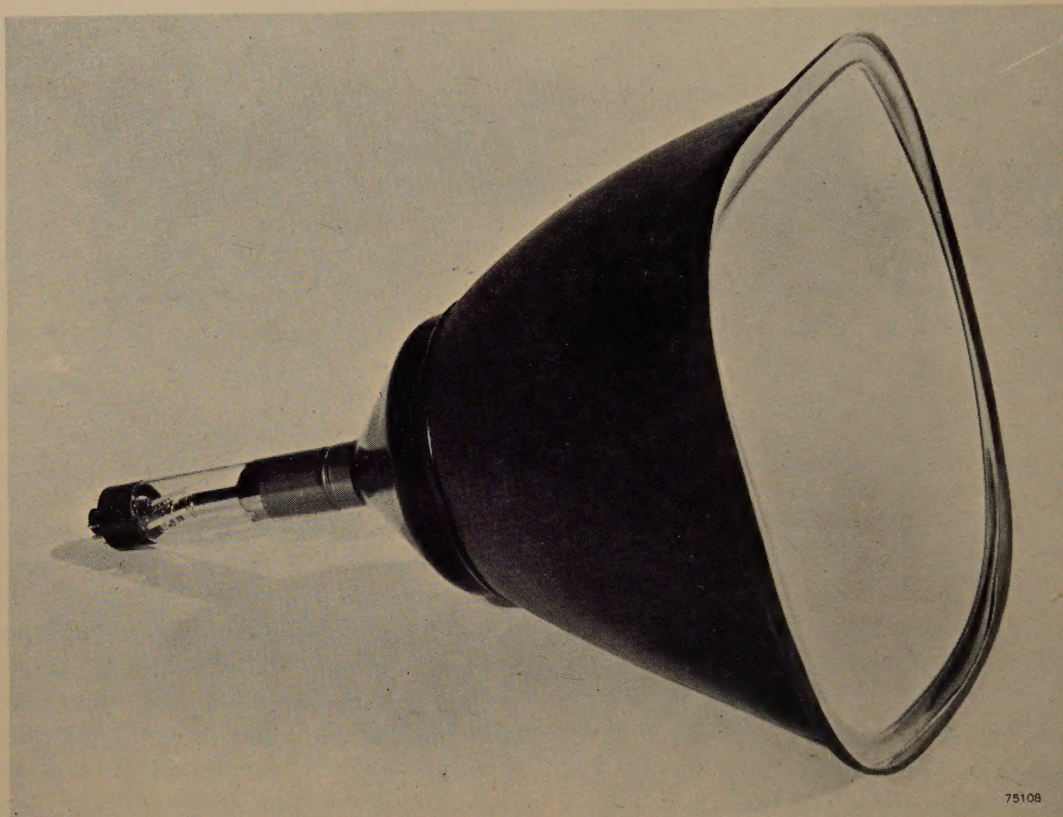

Fig. 4. Picture tube type MW 43-43 with rectangular window (diagonal 43 cm) and metal cone. During the course of 1953 the production will change over from chrome-iron to chromised steel as cone material.

best compromise between thickness and curvature is, generally speaking, not very satisfactory.

2) The larger the tube, the more serious the consequences of a possible implosion, against which safety precautions must be taken.

3) The cost and weight of the tube increase rapidly with increasing dimensions and, in the case of a medium sized tube, form a considerable proportion of the cost of the whole receiver.

It was therefore an interesting development when a tube of 41 cm diameter — which according to the ideas at that time was very large — came on the market in America in 1949, with a cone of chrome-iron. Philips make a similar tube (type MW 43-43) with a rectangular window of diagonal 43 cm and a

1·3 tons exerted by the atmosphere on the 43 cm diagonal window. These figures compare favourably with those of an all-glass tube of comparable dimensions: in such a tube the glass in the middle of the window is 7·5 mm thick and the radius of curvature is about 60 cm, whilst at the edge the glass is much thicker and more strongly curved (*fig. 5*). The whole area of the slightly and evenly curved window of the metal tube can be usefully employed without visible distortion at the corners of the picture.

The fact that the window of the metal tube is only slightly curved has further advantages. Firstly it gives a saving (albeit small) in the total length of the tube and hence in the necessary depth

of cabinet [4]). A second advantage is the following. The high brightness which is obtained with present-day direct view tubes, makes it unnecessary to completely darken the room in which one watches the television; for various reasons this is in itself undesirable. Thus, when one views in the evening, one or more lamps will usually be burning, so that the possibility arises that the viewers see distracting reflections in the tube window. In the case of the almost flat window of a metal picture-tube, it is not very difficult to arrange the lamps, the receiver and the viewers in relation to each other so that the latter no longer see the reflections. It is quite otherwise with the strongly and unequally curved window of a glass picture-tube; distracting reflections are then difficult to avoid.

The special form of the window of a glass tube, dictated by strength considerations, is obtained by pressing. It is, however, difficult to make pressed glass windows for these large tubes completely flawless. The surface possesses a certain roughness, which tends to spoil the sharpness of the image, and can be removed only by polishing.
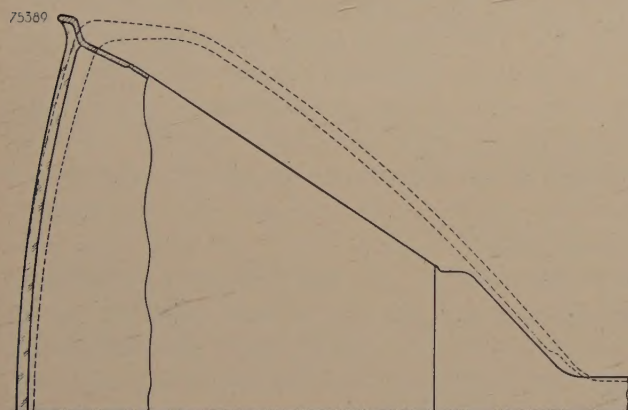
Fig. 5. Full lines: the steel-walled bulb of the tube MW 43-43 in cross-section. Broken lines: cross-section of a wholly glass tube with a window of the same size. In the former the window is less curved, thinner and of equal thickness all over.

This objection does not arise with metal tubes, since windows may be employed which are manufactured by a similar drawing procedure as for ordinary flat window glass and which possess a much smoother surface than windows of pressed glass.

The oldest fully automatic drawing process is that of Fourcault. In this an oval refractory stone — the "débiteuse" — having a slit, rests on the molten glass. The "débiteuse" is pressed downwards, thereby forcing the glass upwards through the slit. At the start of the operation the glass is taken up by a horizontal metal rod which is raised vertically.

The glass, which solidifies to a flat sheet, shows no tendency to constrict, and is thereafter continuously drawn upwards between rollers and repeatedly cut off at a given length.

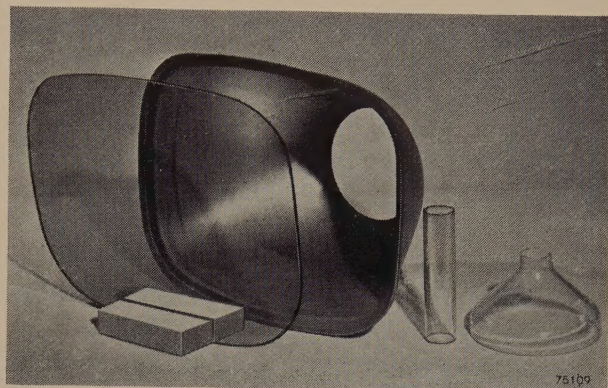Other well-known methods of drawing glass sheet are the Libbey-Owens and the Pittsburgh processes.

Fig. 6. Component parts of the bulb of the picture tube MW 43-43. From left to right: window of grey lead-free "iron" glass, cone of chromised steel, neck and dome of lead-containing "iron" glass.

Panels of the desired shape are cut from the flat sheet (*fig. 6*). These are pressed or "sagged" into a curved form and then sealed to the metal cone. As we shall see later, this is carried out in such a way that implosion of an exhausted bulb seldom, if ever, occurs.

The steel picture-tube

In the above-mentioned tube with chrome-iron cone, the chromium content was an important point of consideration. Chrome-iron with a content of 28% Cr was originally employed. Chromium is a rather expensive metal, however. In addition it is difficult to obtain chrome-iron of uniform composition; local faults are numerous and rejects are relatively high. A third disadvantage of chrome-iron with high Cr-content is its hardness, which renders it less suitable for the radical deformation from a flat sheet to a deep cone.

In our case, as also elsewhere, a change was therefore made to a material with an appreciably lower Cr-content (18%). The risk was thereby introduced of reaching an alloy range in which austenite may appear during the sealing-on of the glass, such an alloy being therefore really chrome-steel. The expansion coefficient of this austenitic chrome-steel is inacceptably high. The danger of austenite formation may, however, be excluded by the introduction of a small proportion of titanium or niobium. In this way a reliable and reasonably workable material [5]) is obtained, which is used on a

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[4]) This point forms the subject of an article to be published shortly in this Review.

[5]) The metallurgical development work has been directed by Dr. Ing. E. M. H. Lips, of the Metallurgical and Plating Laboratory at Eindhoven.

large scale for the manufacture of metal picture-tubes.

Although an important saving in cost was achieved by the reduction of the chromium content from 28% to 18%, this alloy remains nevertheless appreciably higher in price than ordinary mild steel and many more or less special types of steel. Furthermore, most kinds of steel are softer than chrome-iron and thus require less working in the fabrication of a cone from sheet, so that simpler equipment is adequate. With these considerations in mind the development of metal picture-tubes could not yet be considered complete. In the use of steel for picture tubes, two main problems present themselves:

1) A glass matched to steel must be developed.
2) Precautions must be taken to protect the steel, both from chemical attack by substances with which it comes into contact during the manufacture of the tube and from rusting in the course of time. (Chrome-iron is fairly resistant, so that for this material the problem of protection raises no difficulty.)

We shall now discuss the glass and the corrosion questions in turn.

The glass

The "iron" glass developed earlier, as mentioned in the introduction, and used in the tube shown in fig. 3, is less suitable for the windows of picture-tubes for the following reasons:

1) It has a fairly low softening point, so that pumping of the tube at sufficiently high temperatures presents difficulties.
2) It is not easily drawn by the window-glass technique.
3) It contains a considerable proportion of a fairly costly material, namely lead.

These difficulties have now been overcome by the development in the Philips Glass Factory in Eindhoven of a new "iron" glass, which is completely lead-free [6]. The principal characteristics of the old and the new "iron glass" are given in the *table* below.

The composition of the new glass is so chosen that the expansion coefficient of the steel and that of this glass have the same relationship as the expansion coefficient of chrome-iron and that of its appropriate glass. The implosion risk — to which we will refer later — is thus no greater than in chrome-iron tubes. The expansion coefficient of the

glass and that of the steel are regularly checked by sealing samples of these materials to a standard glass, in accordance with a fixed procedure. The strain at the boundary region, measured with the help of polarised light and a compensator [7], is a measure of the difference in expansion coefficients.

We see from the table that the temperatures at which the viscosity reaches certain values (in glass technology comparisons between glasses are always made on this basis), are 40 to 50 °C higher for lead-free glass than for lead-containing glass. This difference is large enough to make good degassing of the tubes possible.

The glass has a considerable chemical resistance. It may be compared in this respect with the glass used for chrome-iron picture-tubes.

When the tube is in use, the fluorescent screen and the cone have a potential of 12-14 kV with respect to earth. The specific resistance of the glass is so great that in spite of this high voltage on the inside of and around the window, no appreciable voltage appears on the outside.

Table. Some characteristics of a lead-containing and a lead-free "iron glass".

	Lead-containing	Lead-free
Expansion coefficient between 30° and 300 °C	11.6×10^{-6} per °C	11.8×10^{-6} per °C
Compression strain after being sealed to mild steel, in mμ/cm *)	350-400	350-400
Lowest annealing temperature **), (°C)	400	450
Highest annealing temperature ***), (°C)	425	470
Softening point †), (°C)	605	645
Temp. (°C) at which specific electr. resistance ††) = 3.33×10^6 Ω cm	370	355
Specific electrical resistance (Ω cm), at 50 °C at 100 °C	3×10^{15} 7×10^{13}	1.5×10^{15} 7×10^{12}

 *) See article referred to in note [7]).
 **) Temperature at which the viscosity reaches the value $10^{14.6}$ poise.
 ***) Temperature at which the viscosity reaches the value $10^{13.6}$ poise.
 †) Temperature at which the viscosity reaches the value $10^{7.6}$ poise.
 ††) Measured as the temperature at which a rod of the glass, 50 mm long and of 5 mm diameter, passes a current of 1.18 μA at a potential difference of 100 V.

[6]) J. Smelt, Le verre pour la télévision, Verres et réfractaires 6, 75-82, 1952 (No. 2).

[7]) A. A. Padmos and J. de Vries, Stresses in glass and their measurement, Philips tech. Rev. 9, 277-284, 1947/1948.

By the use of certain additions the glass can be made grey, i.e. uniformly absorbent in the whole visible spectrum (see fig. 6). In compensation for the loss of light caused by this absorption, the picture loses less contrast due to light coming from outside. This can be explained as follows. To reach the eye of the viewer, the light falling on the fluorescent screen from outside must pass through the window twice, whilst the fluorescent light only passes through once. The stray light is thus attenuated more than the desired light, i.e. the contrast becomes greater as the window absorption increases. A favourable compromise is reached at 35% absorption; the picture is then still bright enough and sufficiently contrasty in a lighted room.

Apart from the window, the tube is composed of the following glass parts: the stem, on which the electron-gun is mounted, the neck, into which this stem is sealed, and the so-called dome, which forms the bridge between neck and cone. The window, cone, neck and dome are shown in fig. 6. A few words now follow about the glass used for the stem, neck and dome.

Even more than in the case of the window it is important that the glass of the dome and neck is of very large specific resistance, as it forms the insulator between the cone, which is at very high tension, and the conducting layer on the inside of the glass on the one hand and the earthed focusing and deflection coils on the other. Although the new, lead-free, "iron" glass is sufficiently insulating to serve for dome and neck, preference is given for these parts to the still better insulating lead-containing "iron" glass, with an eye also on the lower softening point of the latter. This simplifies the manufacture of the stem and also the sealing-in. The softer nature of this glass necessitates, naturally, that during pumping the parts mentioned must be somewhat less strongly heated than the window. This is no disadvantage, as the fluorescent screen is mainly responsible for the gas evolution.

The stem is also made of lead-containing "iron" glass. It is made ring-shaped, so that the distance between the lead-in wires is as large as possible.

Various materials have been tried as lead-in wires for the stem.

Pure iron wire ("Armco"-iron) may be employed provided that care is taken that it is not so exposed to oxidation that a thick, porous oxide-layer is formed. It is advantageous to pre-oxidise in a special atmosphere: in carbon dioxide and steam a dense, hard oxide is formed. To prevent over-oxidation, however, this must be immediately covered with an enamel layer, whereafter the wire must be very carefully handled.

Even less satisfactory are copper-clad iron wire, which develops small cracks around the wire, and chrome-plated iron wire, which displays irregular diffusion phenomena after heat-treatment.

Finally, chromised steel wire and dumet wire (nickel-steel wire in a copper sheath) were found to be very reliable. Chromised material is discussed further in the following paragraph.

Protection against corrosion

During the sealing-in of the window, when localised temperatures between 1050° and 1100 °C are reached, a thick oxide layer would be formed on the steel if the latter was not protected. This skin, loose in structure and porous, would make the the vacuum-tightness of the seal extremely doubtful.

Further risks of attack arise during the introduction of the fluorescent layer, which is allowed to settle from a weakly alkaline liquid [8]. By this treatment, and also during the course of a few days, by the action of the air, loose particles of rust may be produced which, after dropping from the metal, can appear on the screen as dark spots or, still worse, get on to the cathode and impair the emission.

Lengthy trials were necessary to find a satisfactory way of overcoming the risk of corrosion. The following methods have been investigated: enamelling, nickel plating, passivating and chromising. Of these, chromising has led to the most reliable solution. The chromising process [9] is widely used in Germany for making all types of steel articles chemically more resistant. The process results in the formation of a thin layer of chrome-iron on the surface of the steel by diffusion of chromium, thereby supplying the desired corrosion-protection.

The chromising is carried out as follows. The articles to be treated — in our case steel sheets 3 mm thick — are heated with ferro-chrome of very high chrome-content in a closed oven, to about 1000 °C in a hydrogen atmosphere. From time to time a small amount of gaseous HCl is blown through the oven. This forms with ferro-chrome the volatile chromium chloride. The chromium atoms of the volatile compound now undergo an

[8]   See the illustrations in Philips tech. Rev. **9**, 339, 1947/1948 and **10**, 306, 1949.

[9]   A description of this process, which appeared in Germany shortly before 1939, is given by R. Beeker, K. Daeves and F. Steinberg, Oberflächenbehandlung von Stahl durch Chromdiffusion, Stahl u. Eisen **61**, p. 289, 1941; Metallwirtschaft, **20**, p. 217, 1941; and Z. Phys. Chem. A. **187**, p. 354, 1940. See also R. L. Samuel and H. A. Lockington, The protection of metal surfaces by chromium diffusion, Metal Treatment, **18**, Aug.-Dec. 1951 and **19**, Jan.-Febr. 1952.

exchange process with the iron atoms of the steel, which have practically the same size. In this way a layer of chrome-iron is formed on the surface; the layer has a thickness of from 70 to 150 μ, depending on the temperature and time of treatment. The chromium content is about 30% at the outside, decreasing towards the inside. The layer is very intimately bonded to the parent material. The latter can be etched away, however, and *fig.* 7 shows the chrome-iron skins which remain after the steel has been etched away. This test demonstrates at the same time that chrome-iron is more corrosion-resistant than steel.

Fig. 7. Demonstration test of the corrosion resistance of chrome iron produced by chromising steel sheet. A portion of the steel is etched away, whilst the two chrome-iron skins remain unattacked.

The chromised sheets are easily fabricated into a cone, provided that certain precautions are taken and that the steel composition conforms to particular requirements. Thanks to the high Cr-content at the surface, the sealability to glass is very good. Since the chrome-iron layer is very thin, the chromised sheet has an expansion coefficient which hardly differs from that of unclad steel. Nevertheless the layer is sufficiently resistant to repeatedly withstand the treatment which a bulb undergoes

during manufacture; the cone from a rejected bulb can thus be used again, without having to be re-chromised. To protect the exterior of the cone over a long period against rusting, and at the same time to improve the outward appearance, it is lacquered at the end of the manufacturing process.

Chromised steel wire is used for the lead-in wires of the stem. This wire is 0·5 mm diameter, the chrome-iron layer being 15 to 20 μ thick.

Manufacture of bulbs with steel cone

Fig. 8 shows the cone in various stages of manufacture. The forming takes place in two steps. The 3 mm thick plate, cut to the desired size, is first spun into a round cone the height of which is about $5/6$ of that of the final product. The thickness is reduced by this treatment by about one half. The intermediate product is now placed under a heavy press. A punch and die are used to provide the cone at the wide end with an opening of the correct shape and size ("rectangular", with rounded corners and slightly curved sides). At the same time the rim acquires the profile illustrated in fig. 5. In this second treatment, which is more a folding than a drawing process, the material undergoes no reduction in thickness. The rim thus has the same thickness as the original sheet (3 mm), as is required to withstand the heavy mechanical stress present in the rim after exhausting.

After the bottom is punched out and the superfluous portions of the rim are removed, the cone is ready for the sealing-in of the glass parts. The dome is sealed in first (already joined to the neck, into which the stem with the gun will later be sealed), and then the window. In mass production, machines are employed for the processes involved in the sealing-in of the window. Twelve rotating heads are mounted on a slowly revolving table as in the sealing of lamps and radio valves. A cone (with dome and neck) and a window are inserted regularly as a position comes free. Several large flames of gas and air pre-heat these parts successively, and in one of the following positions of the revolving table the actual sealing-in is effected by means of a large number of smaller flames of gas and oxygen.

The sealing-in is completed before the bulb has made a complete revolution of the table. The upper part of the bulb then passes into the so-called equalising oven, the purpose of which is to bring the rim of the cone and the window to the same temperature. This temperature is so chosen that the glass remains somewhat soft.

The bulb leaves the oven after about seven minutes and is then allowed to cool in the air. This

a   *b*   *c*   *d*
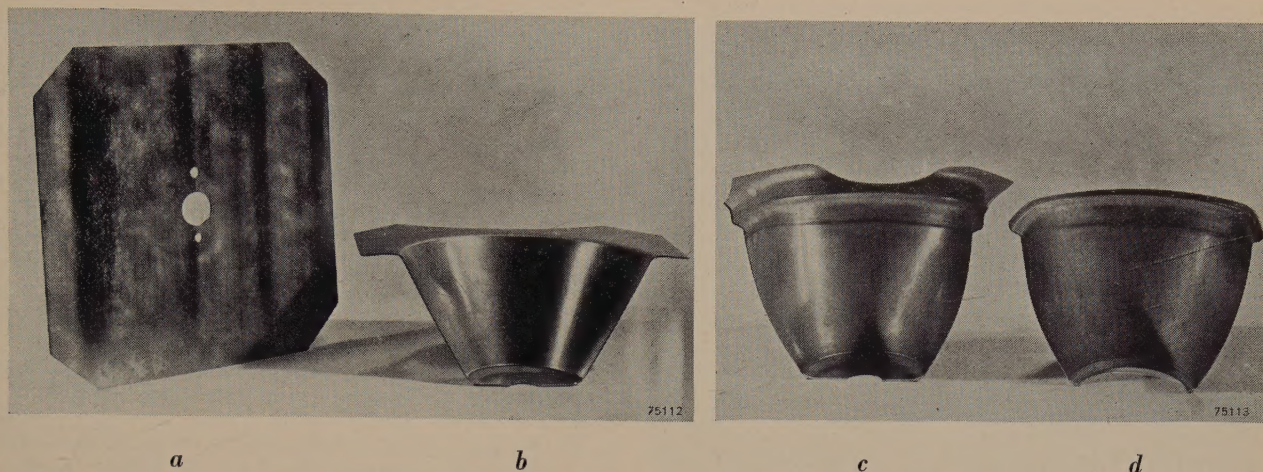
Fig. 8. Forming of a steel cone. *a*) 3 mm thick sheet of chromised steel, cut to the correct size. *b*) This plate is formed by spinning into a circular truncated cone. *c*) The cone has been given the desired shape, with the aid of a punch and die (the cross-section has a 'rectangular' shape with rounded corners and somewhat curved sides), whilst at the same time the rim has been given the correct profile. *d*) Superfluous portions of the edge and bottom have been removed.

differs from the commonly used cooling procedure for most all-glass objects; these remain for a long time in annealing ovens, in which they are brought to room temperature very gradually. There the intention is to achieve a strain-free end-product. In our case, on the other hand, it is in fact favourable if a certain amount of strain remains in the cooled glass, and in particular a compression-strain (see the Introduction).

Strain in the glass

The stress condition which will result in the present case depends on the expansion coefficients a_m and a_g of the metal and the glass respectively, the temperature of the equalising oven and the temperature T_s at which the glass begins to become elastic (i.e. at which stresses begin to develop in the previously soft glass).

As soon as the bulb leaves the oven, the cone and window begin to decrease in temperature; because it is steel, the cone cools off more quickly. When the glass reaches the temperature T_s, therefore, the metal is already at a lower temperature. If $a_m = a_g$, the glass would certainly reach a condition of tension with further cooling, as the circumference of the rim will contract further by a smaller amount than the circumference of the window. Actually, however, $a_m > a_g$ ($a_m = 13 \cdot 0 \times 10^{-6}$ per °C, $a_g = 11 \cdot 8 \times 10^{-6}$ per °C). The possibility thus arises in principle that the glass is under compression in the final condition. The proviso in this connection is that, at the moment at which the glass reaches the temperature T_s, the temperature of the metal shall not have decreased too far below T_s. This stipulation is easier to meet when the temperature

of the equalising oven is not chosen too high above T_s. By regulation of the oven temperature it is thus possible to bring the glass into the desired compression condition.

During the cooling of the glass another effect appears. When the hot window leaves the equalising oven, the outer layers of the glass cool more quickly than the inner ones. The latter will still be plastic at the moment at which the outside begins to become elastic. With further cooling, therefore, the outsides come under compressive forces. This effect ("toughening") gives an extra strength to the glass and is therefore often applied to the windows of motor-cars. The effect is proportionately stronger as the temperature of the equalising oven is made higher, but as we have seen above, the pressure exerted by the cone on the window is then smaller. One is thus forced to a compromise in the choice of equalisation temperature.

To obtain an impression of the strength of the bulb, it can be sealed off at the neck-end and subjected in a pressure vessel to a gradually increasing over-pressure. The bulb window will become flatter to a certain extent and hence attempt to increase its circumference, i.e. the tangential compression-strain originally present becomes smaller and changes over at a certain moment to a tension. With an over-pressure of 3 to 3·5 atm. the window fractures (*fig. 9*). The radial cracks show that the glass has failed due to tangentially operating tensile stresses. The metal is still intact, thanks to its greater tensile strength (~ 2100 kg/cm², compared with that of the glass 300 to 700 kg/cm²). It is for this reason that the bulb has not imploded, i.e. the glass splinters remain in place. Sometimes
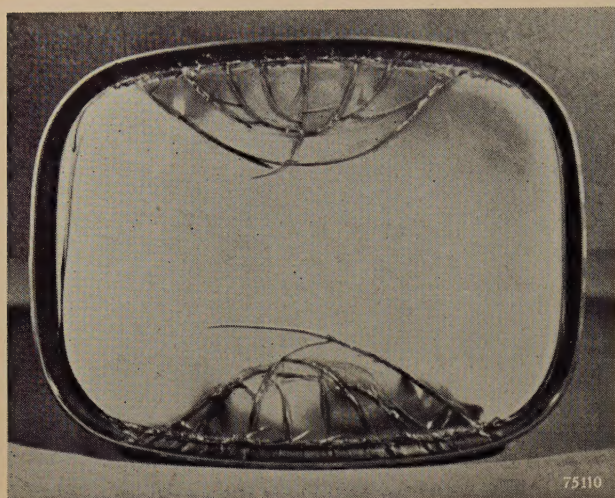
Fig. 9. A sealed-off bulb is exposed to excess pressure, which is increased until the window fractures. The glass does not implode; the splinters, on the contrary, remain in place.

they are pressed so strongly together that a tube with a cracked window can still be fully evacuated.

Investigation of the strain pattern

From the above it will be clear that it is very important to have the correct strain conditions in the window. A check on this is thus indispensable. This is carried out by tests on random samples. A small percentage of the manufactured bulbs are exposed to an over-pressure, as described. A larger percentage is subjected to a non-destructive examination with the help of polarised light. (See, for example, the article referred to in note [7]). This method depends on the fact that mechanical stresses cause strain in the glass and make it double-refracting, and that the extent of the double-refraction is proportional to the applied stresses, or, more precisely, to the difference of the principal stresses. The proportionality-factor depends on the type of glass. The double-refraction can be detected by a change in colour of the transmitted light. The test consists in partially evacuating a bulb. The strains caused by the excess pressure of the atmosphere superimpose themselves on those already present. Thus it is possible to determine the under-pressure Δp_0 in the bulb at which the strains at the edge of the window exactly compensate each other; the colour of the edge area then changes over. The criterion for the assessment is that Δp_0 must be between two empirically determined limits.

Further description of the strain examination

A polarised beam of light — obtained by use of a "Polaroid" filter — passes through the bulb; it enters through the window and goes out through the dome. It then passes through a so-called red plate [10]) and a second "Polaroid" filter. This latter transmits only light whose vibration direction lies perpendicular to that of the former. The beam passing through the second filter is observed.

The principal stresses in the glass vary through the thickness at the point considered, as a consequence of the curvature of the window. In *fig. 10a* this is depicted for one of the principal stresses, in this case for the tangentially directed principal stress σ_1. For the radially directed principal stress σ_2 (perpendicular to the plane of the illustration) a similar condition arises; the third principal stress, perpendicular to σ_1 and σ_2, is negligibly small. The variation of σ_1 in the thickness of the window can be represented by the superposing of a uniform compression stress (or it might be a tensile stress), σ_{10} (fig. 10b), and a symmetrically distributed bending stress (fig. 10c). The same applies to the radial stress σ_2. The bending stress does not contribute to the optical effect since the double refraction caused by the positive bending stresses is exactly compensated by an opposing double refraction caused by the negative bending stresses. The resultant double refraction at a point in the window is thus a measure only of $\sigma_{10}-\sigma_{20}$ at that place, and this applies also in good approximation to the total double refraction (of window and dome together), since the glass of the dome is practically strain-free and thus causes no double refraction worth mentioning. It should be noted that the method gives no definite information about σ_{10} or σ_{20} independently, but only about the **difference** between these two principal stresses.

This stress difference $\sigma_{10}-\sigma_{20}$, depends on the distance r of the point under consideration from the centre of the window. Let us assume that the window is circular, as was the case with older tubes. It follows then from the symmetry that in the centre σ_{10} is equal to σ_{20}. No double refraction appears, therefore, i.e. no colour change is to be seen in the centre of the window. Outside of the centre colour variation is observed, in proportion to the magnitude of the local strain-difference.

The strain pattern in the window of an evacuated bulb is due to the superposition of the straining imposed during the sealing of the window and its subsequent heat-treatment and of the strain which arises from the over-pressure of the outside air. These stresses can be separately calculated for the rotationally symmetrical case described here [11]); for a rectangular tube this is not possible. Thus the curve *l* in *fig. 11* gives the calculated variation of $\sigma_{10}-\sigma_{20}$ as a function of r, in so
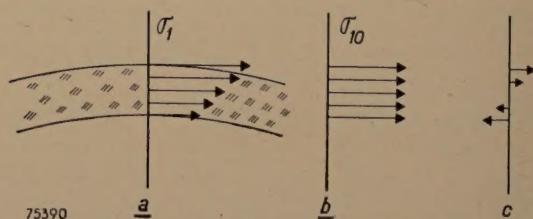

Fig. 10. *a*) Cross-section of the window of a metal picture tube, showing the variation of the tangentially directed principal stress σ_1 along the perpendicular at a given point. This variation can be represented by the superposition of an evenly divided compression (or tension) stress σ_{10} (see *b*) and a symmetrically divided bending stress (see *c*).

[10]) The function of the red plate is as follows. Without this plate the interference colours produced by double refraction in the glass are grey-blue in tint and therefore difficult to interpret. The red plate gives an appreciable, constant, double refraction corresponding to the interference colour "red of the first order". Changes in this colour as a consequence of double refraction in the object examined are easy to interpret.

[11]) The calculations were carried out as suggested by Dr. Ir. J. A. Haringx of the Philips Research Laboratories, Eindhoven.

far as the stresses are caused by the difference between the expansion coefficient of the metal (α_m) and that of the glass (α_g). The curves 2 give the calculated values of $\sigma_{10}-\sigma_{20}$ in so far as these stresses are caused by the difference in air pressure; each curve corresponds to a certain pressure difference Δp. Superposing the curve 1 and one of the curves 2 gives the curve 3, which shows the total difference in principal stress at a certain pressure difference Δp. The magnitude of Δp can be so chosen in the test described that 3 passes through zero when r is equal to the radius R of the window, as represented in fig. 11. In this condition the glass shows no double refraction at the edge.

In this way we arrive at an empirical check on whether the sealed-in windows have the correct strain pattern. If air is pumped out of the tube until the colour of the edge area is
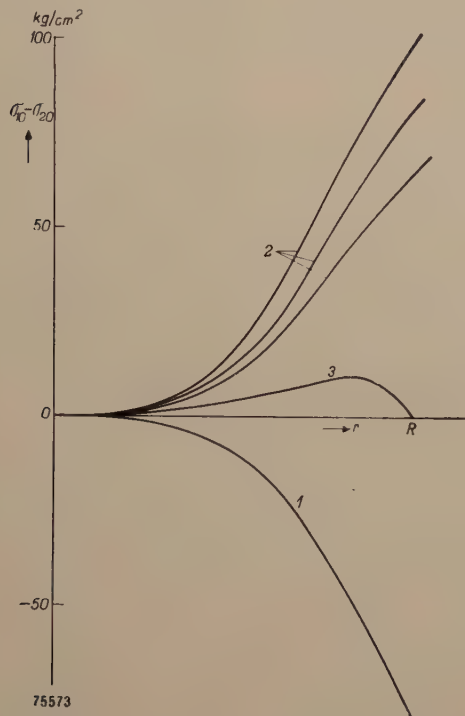
seen to change over, then that curve has been chosen from the group of curves 2 which the curve 1 compensates, at least at the edge. The pressure difference Δp_0 at which this occurs is a measure of $\sigma_{10}-\sigma_{20}$ at the edge. This difference in pressure
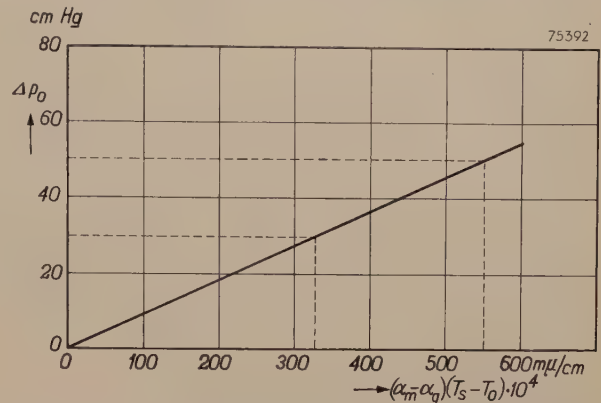
Fig. 12. The excess pressure Δp_0 which is necessary to make $\sigma_{10}=\sigma_{20}$ at the edge of the window, is a linear function of the difference in circumference $(\alpha_m-\alpha_g)(T_s-T_0)$, which the cone rim and the window would show if they were cooled down separately from T_s to T_0.

is proportional to the amount of pre-straining in the window, i.e. to the relative difference in linear dimensions that the window and the cone rim would show if they had been allowed to cool down separately from the softening temperature T_s of the glass to the room temperature T_0. The pressure difference Δp_0 is thus proportional to $(\alpha_m-\alpha_g)(T_s-T_0)$. This linear relationship [12]) is shown in fig. 12, as calculated for circular tubes.

The calculation cannot be carried out for rectangular tubes, but a similar relationship holds and the same test can be made. In this case, the test is applied near the middle of the longest edge, this being the weakest spot. For rectangular, as for round tubes, the criterion is that Δp_0 must be between certain empirically established limits.

Fig. 11. The variation of $\sigma_{10}-\sigma_{20}$, calculated from the expansion coefficients α_m and α_g (curve 1), and the calculated values of $\sigma_{10}-\sigma_{20}$ resulting from various differential air pressures Δp, (curves 2), as functions of the distance r from the centre of a circular window (radius R). Curve 3 is obtained by superposition of curve 1 and one of the curves 2, so selected that 3 goes through zero at $r = R$.

[12]) As described on p. 282 of the article referred to in note [7]), the expansion coefficient of a material can be very accurately measured by means of a strain examination (with the help of polarised light) after sealing to a plate of a standard material.
The amount of the observed double refraction per cm thickness of the material is expressed in millimicrons wavelength difference in the light rays. For this reason the abscissae of fig. 12 and the ordinates of fig. 2 are provided with a scale in mμ/cm.

Summary. One of the ways in which the demand for larger television images has been met, is the construction of larger picture-tubes for direct viewing. An important step in this development was the change-over from all-glass tubes to tubes with a metal conical portion. The window can then be drawn according to window-glass technique. A much more even thickness can be obtained by this procedure than by glass pressing, which is applied for the production of the thicker, stronger curved window of all-glass tubes. This results in an improvement in picture quality. Thanks to the reduced curvature, the whole window surface is usefully employed and reflections from lamps in the room are easily avoided.

In the most recent development the cone metal is made of mild steel, with which useful experience in similar applications had already been gained. It is considerably cheaper and more easily worked than the chrome-iron used hitherto. A new type of glass had to be developed, however, which, while having an expansion coefficient near to that of steel, had also a sufficiently high softening point (with a view to

the de-gassing of the tube.) Success has been achieved in making such a glass (lowest and highest annealing temperatures 450 and 470 °C respectively, softening point 645 °C). This glass is lead-free, which is favourable as far as price is concerned. Furthermore, it can be made grey to reduce loss of contrast caused by light coming from outside. To protect the steel from corrosion during fabrication of the tube, the chromising process was finally adopted, in which a layer of chrome-iron is formed on the surface. Chromised steel is also used for the lead-in wires of the stem. The sealing of the window into the cone and the subsequent cooling are carried out in such a manner that in the final condition of the cone the edge of the glass is put under pressure and at the same time compression strain is set up in the outermost layers of the glass ("toughening"). The window owes its great strength, in spite of its small thickness, to this fact. The risk of implosion is appreciably smaller than with glass tubes of the same dimensions. The achievement of the correct strain pattern in the glass is checked with the aid of polarised light.

MECHANISM OF THE SYNCHRONIZATION OF *LC*-OSCILLATORS

by J. VAN SLOOTEN.   621.396.018

Electric oscillators whose frequency is governed by an inductance and a capacitance show changes in phase and frequency when influenced by an interference, the best known phenomenon being synchronization, where a periodic interference of sufficient strength causes the oscillator to vibrate in its own frequency. If the interference is too weak to synchronize the oscillator, a beat phenomenon arises in combination with a small variation of the mean frequency. Assuming the oscillator voltage to be practically sinusoidal, the author proceeds to deal with these phenom- ena with a view to giving a clear picture of their mechanism. With the method followed it is also possible to investigate the frequency fluctuations under the influence of a non-periodic disturbance, for instance the thermal noise voltage of the oscillator circuit.

Introduction

In radio engineering use is made, on a large scale, of electric oscillators which produce an approximate- ly sinusoidal voltage. Mostly these consist of an oscillatory circuit formed by an inductance and a capacitance in combination with a triode feeding the circuit with a current which is in phase with the alternating voltage across the circuit. In a certain sense the triode then functions as a negative resis- tance shunted across the oscillator circuit. As a rule the amplitude is limited by a shifting of the working point towards a comparatively high nega- tive grid bias as the amplitude increases. This takes place automatically by the rectifying action of the diode formed by the grid and cathode. The properties of these technical oscillators have already been discussed at length in this journal [1]), where it was made clear that the manner in which the amplitude is limited influences greatly those proper- ties of the oscillator which are related to the mean amplitude, as for instance the stability of operation. There are, however, also properties of the oscillator which, in the first instance, are independent of the amplitude-limiting mechanism. Among these is the susceptibility of the phase and frequency to extraneous disturbances, an example of which is the known property of an *LC* oscillator [2]) being s y n c h r o n i z e d by an external periodic signal of sufficient strength and of a frequency differing from the natural frequency of the oscillator.

About this synchronization phenomenon a number of papers has already been published, almost entirely based upon the original investigations of Möller, Appleton and, particularly, of Van der

Pol [3]). These publications are characterized by very general mathematical reasoning, so that also those oscillators whose damping and regeneration are so great as to cause the mode of oscillation to deviate greatly from a sine curve (relaxation oscillations) have been included in the investigations. When a f t e r w a r d s introducing into the results the assumption that the mode of oscillation is practi- cally sinusoidal, it is found that there is a very simple condition for synchronization [4]). A disad- vantage attaching to this method is that the me- chanism of the synchronizing process is not made clear. When, however, one starts right from the b e g i n n i n g by assuming the mode of oscillation to be practically sinusoidal, it appears that an entirely different method of treating the problem is possible, a method which is not only simple but also has the advantage that no assumptions need be made about the way in which the amplitude of oscillation is limited. From the following description of this method it will be seen that in this way it is also possible to give an elementary treatment of the beat phenomenon occurring when the interfering signal is too weak to obtain synchronism. It is then only a step farther to investigate the irregular frequency fluctuations arising in the case of an arbitrary disturbance, as for instance a source of noise in the oscillator itself.

Phase variation under the influence of one single short pulse

Fig. 1 represents a common type of oscillator, where the interfering signal is assumed to act as a

[1]) J. van Slooten, Philips tech. Rev. **7**, 40-45 and 171-177, 1942.
[2]) For the sake of simplicity the term *LC* oscillator is used for an oscillator whose frequency is governed by an inductance *L* and a capacitance *C*.

[3]) H. G. Möller, Jahrbuch drahtl. Tel. **17**, 256-287, 1921.
    E. V. Appleton, Proc. Cambr. Phil. Soc. **21**, 231-248, 1922.
    B. van der Pol, Phil. Mag. **3**, 65-80, 1927.
    B. van der Pol, Proc. Inst. Rad. Engrs. **22**, 1051-1086, 1934.
[4]) G. Francini, Alta Frequenza **18**, 125-133, 1949.

source of current shunted across the *LC* circuit, which is represented by the pentode at the left, and the interfering voltage *s* applied to its control grid.
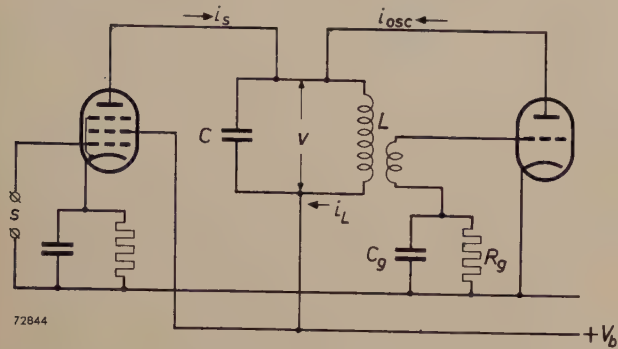

Fig. 1. A conventional oscillator circuit. The disturbing voltage *s* acts upon the grid of the pentode at the left.

The amplitude of the oscillation is limited by means of the grid capacitor C_g and the grid resistor R_g, but in the following considerations the nature of the limiting mechanism does not play any part, so that some other method of limiting, for instance by the curvature of the i_a-v_g characteristic, could equally well be assumed.

The line of reasoning is that the oscillator is to be regarded as a freely oscillating *LC* circuit whose amplitude is maintained at a constant level by the oscillator valve feeding the circuit with a current which is in phase with the circuit voltage *v*. Thus that valve influences the amplitude but not the phase of the voltage across the oscillatory circuit. In accordance with this supposition is the result, which will presently be found, that a disturbing pulse, coinciding with a voltage maximum or minimum of the circuit voltage *v*, does not affect the phase of the oscillation.

For the purpose of this investigation the effect is considered of a short pulse fed from the pentode in fig. 1 to the circuit at an instant separated from a preceding voltage maximum by a phase angle φ. This is represented in *fig. 2*.

Let the charge $i_s \Delta t$ given to the circuit by the interfering pulse be :

$$q_t = i_s \Delta t. \quad . \quad . \quad . \quad . \quad . \quad (1)$$

The current i_L through the coil being continuous, the interfering pulse causes a momentary rise in voltage Δv at the capacitor *C*, given by

$$\Delta v = \frac{q_t}{C}. \quad . \quad . \quad . \quad . \quad (2)$$

We shall now make the capacitor voltage v' and the coil current i_L', immediately following upon

the pulse, correspond to that of an undisturbed oscillation, the phase and amplitude of the voltage maximum of which differ from those previous to the pulse. Hence:

$$\left. \begin{array}{l} V_{\max} \cos \varphi + \Delta v = V'_{\max} \cos \varphi', \\ I_{L\max} \sin \varphi = I_{L'\max} \sin \varphi'. \end{array} \right\} \quad . \quad (3)$$

It is, moreover, self-evident that

$$\frac{V'_{\max}}{I_{L'\max}} = \frac{V_{\max}}{I_{L\max}}.$$

Let us now put:

$$\left. \begin{array}{l} V'_{\max} = V_{\max} + \Delta V_{\max}, \\ I_{L'\max} = I_{L\max} + \Delta I_{L\max}, \\ \varphi' = \varphi + \Delta\varphi. \end{array} \right\} \quad . \quad . \quad (4)$$

Since $\Delta\varphi$ is a small angle, we may write: $\cos \Delta\varphi = 1$ and $\sin \Delta\varphi = \Delta\varphi$. Employing a known trigonometric relation, we may then introduce in (3):

$$\left. \begin{array}{l} \cos \varphi' = \cos \varphi - \Delta\varphi \sin \varphi, \\ \sin \varphi' = \sin \varphi + \Delta\varphi \cos \varphi. \end{array} \right\} \quad . \quad . \quad (5)$$

Ignoring the terms containing the products $\Delta V_{\max}\Delta\varphi$ and $\Delta I_{L\max}\Delta\varphi$, it folllows from (3), (4) and (5) that:

$$\left. \begin{array}{l} \Delta V_{\max} \cos \varphi - V_{\max} \Delta\varphi \sin \varphi = \Delta v, \\ \Delta I_{L\max} \sin \varphi + I_{L\max}\Delta\varphi \cos \varphi = 0. \end{array} \right\} \quad (6)$$
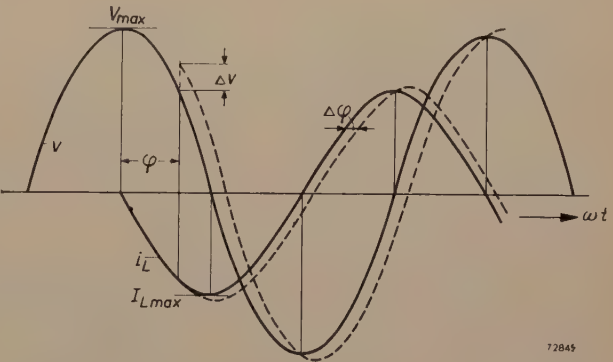

Fig. 2. Diagrammatic representation of the influence of a short interfering pulse upon the momentary amplitude and phase of the oscillator voltage. The phase discontinuity $\Delta\varphi$ is to be particularly noted.

Introducing the relative pulse strength ε and the relative amplitude variation *x*:

$$\varepsilon = \frac{\Delta v}{V_{\max}} = \frac{i_s \Delta t}{C V_{\max}}, \quad . \quad . \quad . \quad . \quad (7)$$

and

$$x = \frac{\Delta V_{\max}}{V_{\max}} = \frac{\Delta I_{L\max}}{I_{L\max}}, \quad . \quad . \quad . \quad (8)$$

eq. (6) can be written:

$$x \cos \varphi - \varDelta\varphi \sin \varphi = \varepsilon, \atop x \sin \varphi + \varDelta\varphi \cos \varphi = 0,\Big\} \quad \cdots \quad (9)$$

from which follows:

$$\boxed{\begin{aligned} x &= \varepsilon \cos \varphi, \\ \varDelta\varphi &= -\varepsilon \sin \varphi. \end{aligned}} \quad \cdots \quad (10)$$

Our further investigation is based upon these equations (10). It is seen that interfering pulses occurring at a voltage peak ($\varphi = 0$ or $\varphi = \pi$) influence the amplitude but not the phase, whereas those occurring when the circuit voltage passes through zero ($\varphi = \pi/2$ or $\varphi = 3\pi/2$) cause only a variation in phase. Further it should be recognized that the amplitude changes are neutralized owing to the limiting mechanism in course of time, but the phase changes remain, or, in other words, it might be said that the oscillator has an "infinite memory" for phase disturbances.

Synchronization by a series of equidistant pulses

The second of the two equations (10) already provides the possibility of synchronization by an external signal, and at the same time it shows the mechanism: when at each cycle of the oscillator voltage an interfering pulse occurs at the phase angle φ, there is a phase discontinuity $\varDelta\varphi = -\varepsilon \sin \varphi$ per cycle, so that there is a relative frequency variation $\varDelta\varphi/2\pi$. The frequency is increased when the pulses occur prior to the voltage maximum and decreased when they occur after that maximum, as is the case in the position represented in fig. 2.

The resulting relative frequency variation is greatest when the pulses lead or lag by an angle $\pi/2$ with respect to the voltage maximum. Pulses of the relative strength ε can cause at most a relative frequency change $\varepsilon/2\pi$. If the relative difference between the pulse frequency and the undisturbed oscillator frequency is greater than $\varepsilon/2\pi$, no synchronization can take place and a beat phenomenon arises, which will be reverted to later. If the relative frequency difference is less than $\varepsilon/2\pi$, synchronization can take place at a smaller phase angle ($-\pi/2 < \varphi < \pi/2$). As long as the state of synchronism has not yet been reached, the phase φ changes from cycle to cycle until the pulses occur at the right value of φ. Experience teaches that once this is reached the state of synchronism continues, even when there are small variations in the strength or the instant of application of the pulses. It would lead us too far to give here the proof of this stability.

From what has been said it is seen that the synchronization of an LC oscillator by means of pulses

can be represented almost as readily as that of a multivibrator or a blocking oscillator. It could be represented by a synonymous picture of a clock being put back 5 minutes after every 65 minutes, so that it runs synchronously with an interference having a cyclic duration of 65 minutes; as a result the clock shows "hours" of 65 minutes.

Synchronization by a sinusoidal voltage

If the synchronizing signal is not pulse-shaped the phase variation per cycle can be found by integrating over a cycle with the aid of the second equation (10), the signal then being regarded as consisting of elementary pulses. Owing to the factor $\sin \varphi$ in (10), this integration corresponds to a determination of the fundamental component of the disturbing signal. It may therefore be concluded that the synchronizing action of a periodically repeated pulse is equal to that of a sinusoidal signal of the same strength as the fundamental component of the series of pulses. The synchronization by a sinusoidal signal can thus be derived in a simple way from that by pulses.
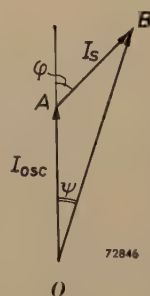

Fig. 3. The vector OA, which is in phase with the circuit voltage, represents the anode alternating current of the oscillator valve. The vector AB represents the synchronizing sinusoidal interfering current. The angle φ is the phase angle between the latter and the circuit voltage, whilst the angle ψ is a measure for the frequency deviation of the oscillator.

The case of synchronization by a sinusoidal signal can also be deduced from the vector diagram of *fig. 3*. The alternating anode current i_{osc} of the oscillator valve is in phase with the circuit voltage. The signal current i_s is not in phase with it. The vector OB represents the total current fed to the LC circuit. The constant phase angle φ between the latter and the circuit voltage is only possible if the oscillator frequency differs from the resonant frequency of the circuit. The diagram does not show how the synchronization is achieved; it only represents the possible final state [5]).

Differential equation for the total phase deviation under the influence of a sinusoidal disturbance

In order to investigate the beat phenomenon arising when the interfering signal is too weak to bring about synchronization, a differential equation will be derived for the general case of a sinusoidal interference whose angular frequency differs by an amount $\varDelta\omega$ from that of the undisturbed oscillator.

[5]) An article on synchronization and the beat phenomenon, in which this diagram is dealt with, was given by R. Adler, Proc. Inst. Rad. Engrs. **34**, 351-357, 1946.

The phase deviation of the oscillator voltage under the influence of the interference is called Φ, which must be considered as a function of time.

The oscillator voltage is therefore assumed to be

$$v = V_{max} \cos (\omega_0 t + \Phi), \quad \ldots \quad (11)$$

and the interfering current fed to the LC circuit is put:

$$i_s = I_{s\,max} \cos (\omega_0 t - \Delta\omega t - \varphi) \quad . \quad (12)$$

The term $\Delta\omega$ is small compared with ω_0, whilst φ is a constant phase angle as yet undetermined. The momentary phase difference between the interference and the oscillator voltage is clearly

$$\Theta = \Delta\omega t + \varphi + \Phi. \quad \ldots \ldots \quad (13)$$

The interfering current is split up into a component which is in phase with the oscillator voltage and one which shows a phase difference of $\pm\pi/2$. The first component influences the amplitude but not the phase of the oscillator, while the reverse applies for the second component. By employing a known trigonometric relation we find from (12) that:

$$
\begin{aligned}
i_s = {} & I_{s\,max} \cos (\omega_0 t + \Phi) \cos (\Delta\omega t + \varphi + \Phi) + \\
& + I_{s\,max} \sin (\omega_0 t + \Phi) \sin (\Delta\omega t + \varphi + \Phi). \quad (14)
\end{aligned}
$$

Here the last product represents the interfering component which is in quadrature with the momentary oscillator voltage and thus reponsible for the phase variations. It is only this component that will now be considered.

In order to arrive at the differential equation referred to, the momentary interfering current has to be multiplied, in accordance with the second equation (10), by the sine of the phase angle through which the oscillator voltage has passed subsequent to the last voltage maximum. From (11) it appears that, to within a multiple of 2π, this angle is equal to $\omega_0 t + \Phi$. Hence, from (7), (10) and (14):

$$
\begin{aligned}
\mathrm{d}\Phi = {} & -\frac{I_{s\,max}\,\mathrm{d}t}{CV_{max}} \sin^2 (\omega_0 t + \Phi) \sin (\Delta\omega t + \varphi + \Phi) = \\
= {} & \frac{I_{s\,max}\,\mathrm{d}t}{2CV_{max}} [1 - \cos 2\,(\omega_0 t + \Phi)] \sin (\Delta\omega t + \varphi + \Phi).
\end{aligned}
$$

From this equation it is seen that the phase deviation Φ is subject to a gradual variation relating to the frequency difference $\Delta\omega$, but that Φ also shows a rapid variation within each cycle of the oscillator voltage. This rapid variation thus bears more the character of a variation of the wave form of the oscillator voltage, rather than that of a real phase fluctuation. The corresponding effect can therefore be ignored, so that finally we arrive

at the following differential equation:

$$\frac{\mathrm{d}\Phi}{\mathrm{d}t} = -\frac{I_{s\,max}}{2CV_{max}} \sin (\Delta\omega t + \varphi + \Phi). \quad . \quad (15)$$

A closer consideration reveals that the rapid phase variation within each cycle has indeed no physical significance but is only related to the fact that we have taken as the undisturbed phase that of a freely oscillating circuit, and not that of a circuit driven by a sinusoidal current with the frequency ω_0. Thus the term ignored was only a complication due to the simplified treatment of the problem.

The case of synchronization is easily deduced from the differential equation (15), for then it is obvious that

$$\Phi = -\Delta\omega t.$$

Substituting this in (15) we find:

$$\frac{I_{s\,max} \sin \varphi}{2CV_{max} \Delta\omega} = 1, \quad \ldots \ldots \quad (16)$$

where, in accordance with fig. 2 and fig. 3, φ represents the phase angle between the interfering current and the oscillator voltage. The condition for synchronization is clearly:

$$\frac{I_{s\,max}}{2CV_{max} \Delta\omega} \geq 1. \quad \ldots \ldots \quad (17)$$

The beat phenomenon in the case of a very weak interference; comparison with a normal beat

When the interference is very weak the phase of the oscillator will deviate very little from that in the case of no interference. An approximate solution of (15) is then obtained by ignoring the term Φ on the right-hand side, thus giving as a first approximation for a weak interference:

$$\Phi = \frac{I_{s\,max}}{2CV_{max} \Delta\omega} \cos(\Delta\omega t + \varphi). \quad . \quad (18)$$

This equation represents a phase modulation with the angular frequency $\Delta\omega$, the difference between the angular frequency of the signal and that of the oscillator. When compared with (17) it shows that the maximum phase deviation must remain below 1 radian, as otherwise (18) is certainly no longer valid as a solution of (15). The exact solution of (15) will be reverted to later.

It is very instructive to compare the beat phenomenon represented by (18) with a normal beat between two fixed frequencies. This is done by replacing the oscillator current i_{osc} in fig. 1 by an equally strong current with the constant frequency ω_0. This gives across the LC circuit a voltage of the amplitude

$$V_{max} = I_{osc\,max} R, \quad \ldots \ldots \quad (19)$$

where R is the parallel resistance of the circuit at resonance. With the series resistance r this bears the known relation:

$$R = \frac{L}{Cr}. \quad \ldots \ldots \quad (20)$$

The interfering current represented by (12) gives, across the LC circuit, the voltage amplitude

$$V_{s\,max} = I_{s\,max} \frac{R}{\sqrt{1+y^2}}, \quad \ldots \quad (21)$$

where y has the value

$$y = \frac{2L}{r} \Delta\omega = 2\,RC\,\Delta\omega \quad \ldots \ldots \quad (22)$$

The sum of the voltage vectors given by (19) and (21) can be found in the known way from a diagram in which $V_{s\,max}$ has the relative angular velocity $\Delta\omega$ with respect to V_{max} (fig. 4).
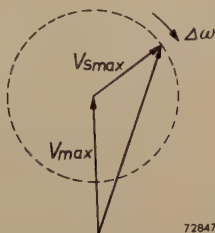

Fig. 4. Vector diagram of a normal beat. With respect to V_{max} the vector $V_{s\,max}$ has the relative angular velocity $\Delta\omega$.

If $V_{s\,max}$ is sufficiently small compared with V_{max} it is clear that, compared with the case where there is no interference, for the normal beat the maximum phase deviation is

$$\Phi_{normal} = \frac{V_{s\,max}}{V_{max}}. \quad \ldots \ldots \quad (23)$$

Finally, from (18) to (23) inclusive, we find for the proportion of the maximum phase deviations for the oscillator and for the normal beat:

$$\frac{\Phi_{osc}}{\Phi_{normal}} = \frac{\sqrt{1+y^2}}{2CR\Delta\omega} = \frac{\sqrt{1+y^2}}{y}. \quad \ldots \quad (24)$$

This shows that, particularly for small values of $\Delta\omega$ ($y<1$), the phase deviation of an oscillator is much greater than that of a normal beat. This phenomenon can be further explained in the following way.

When a disturbance of short duration is applied to an LC circuit fed by a current of constant frequency, a change takes place in the momentary phase. After a time this change decays owing to the circuit damping. In the case of an oscillator, however, this phase disturbance continues to exist for an indefinite time. Consequently (below the limit set by the occurrence of synchronization) in the case of a beat whose cycle is longer than the decay time $2L/r$ of the circuit the phase deviation of an oscillator may become much larger than in the case of a normal beat. This is undoubtedly a remarkable and little known property of an oscillator.

Exact solution of the differential equation (15)

Although a fairly good general picture has already been formed, it has not yet been made clear how in the case of an interference growing in strength the beat phenomenon is transformed into the state of synchronization. The answer to this question lies in the exact solution of (15). It will appear that the beat phenomenon itself is already accompanied by an initially very small change in the mean oscillator frequency.

At first sight solving the differential equation (15) appears to be rather complicated, but it can nevertheless be accomplished with elementary means. For this purpose we introduce as a new variable:

$$u = \Delta\omega t + \varphi + \Phi, \quad \ldots \ldots \quad (25)$$

and for the sake of brevity we furthermore put:

$$\frac{I_{s\,max}}{2CV_{max}\Delta\omega} = a \quad (a<1), \quad \ldots \quad (26)$$

so that (15) becomes:

$$\frac{du}{1 - a\,\sin u} = \Delta\omega\,dt.$$

The left-hand member of this equation leads to an elementary trigonometric integral, by means of which we find as the solution of (15):

$$\frac{2}{\sqrt{1-a^2}} \arctan\left[\sqrt{\frac{1+a}{1-a}} \tan\left(\frac{u}{2} - \frac{\pi}{4}\right)\right] = \Delta\omega \cdot t + \text{const.}$$

By putting $\Phi = 0$ at a suitably chosen moment, and by employing eq. (25), this equation can, without loss of generality, be written:

$$\tan\left(\frac{\Delta\omega t}{2} + \frac{\Phi}{2}\right) = \sqrt{\frac{1-a}{1+a}} \tan\left(\frac{\sqrt{1-a^2}}{2} \Delta\omega t\right). \quad (27)$$

This in turn can be abbreviated to read:

$$\tan\psi_2 = \sqrt{\frac{1-a}{1+a}} \tan\psi_1, \quad \ldots \ldots \quad (28)$$

where ψ_1 and ψ_2 stand for the respective terms in brackets in (27).

From (28) it follows that tan ψ_1 and tan ψ_2 assume simultaneously the value 0 and likewise simultaneously the value ∞. Thus the angles ψ_1 and ψ_2 reach the values 0, $\pi/2$, π, $3\pi/2$ at the same time. This is graphically represented in *fig. 5*. On an average the angles ψ_1 and ψ_2 obviously increase at the same rate. Therefore, disregarding the periodic fluctuation, the average increase of Φ is given by:

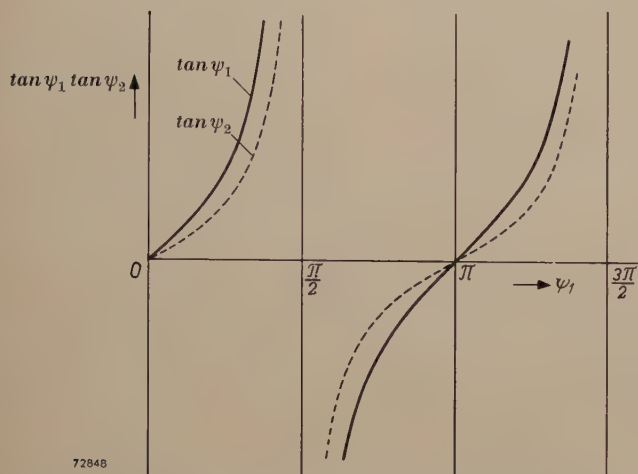$$\Phi = -\Delta\omega t\,(1-\sqrt{1-a^2}).$$

Fig. 5. Graphic representation of the equations (28) and (27). The angles ψ_1 and ψ_2 increase, on an average, at the same rate.

and the average angular frequency difference between the interference and the oscillator is obviously:

$$\Delta\omega\sqrt{1-a^2}.$$

This is in agreement with the known phenomenon that for a not too weak interference the beat note is perceptibly lower than the difference $\Delta\omega/2\pi$ between the frequencies of the interference and the undisturbed oscillator. Thus the mean oscillator frequency is "pulled" by the interference before there is a possibility of synchronization.

From (28) and fig. 5 it can moreover be deduced that, when a approaches unity, the maximum difference between ψ_2 and ψ_1 approaches $\pi/2$. Since the fluctuation in the increase of ψ_2 corresponds to that of $\Phi/2$, it is seen that, as the possibility of synchronization is approached ($a\to1$), the phase fluctuation of the oscillator due to the decreasing beat frequency becomes very slow and approaches the maximum amplitude $\pm\pi$. Thus the oscillator voltage passes through each phase relation with respect to the interference, so that as a increases further, the state of synchronization can come about without any discontinuity. This completes the picture of the phenomena occurring.

Non-linearity of the phenomena; application to arbitrary disturbances; the frequency noise of oscillators

Considering that in the foregoing the non-linearity of the amplitude-limiting mechanism does not play any particular part, it might be thought that a linear treatment of the interference problem has been given. This, however, is by no means the case, as already follows from the unmistakable non-linearity of the differential equation (15).

It is true that the approximative solution (18) is linear, so that it can be applied to a sum of two or more sinusoidal interferences, provided that for each interference the value of a defined by (26) is sufficiently small with respect to unity. If this condition is not satisfied, then the effect of an interference is no longer independent of the presence of other interferences. This is most evident for the case of synchronization: synchronization by two periodic interferences of different frequencies is a *contradictio in terminis*.

In the case of some non-periodic disturbance, for instance a source of noise in the oscillator itself, it might be suggested to take the Fourier spectrum of that disturbance and to apply to its components the linear approximate solution (18). This, however, is not permissible, since as a rule the spectrum will contain components for which the condition $a \ll 1$ is not satisfied. And these happen to be the most interesting components, namely those which are grouped immediately around the undisturbed oscillator frequency [6]).

Another method, and in our opinion this is the only correct one, is to start from the time function of the disturbance, thus in the case of a noise source from the real statistical fluctuations. One then arrives at a mean value of the phase fluctuation per cycle. This will not be gone into further here because we hope to revert to this subject elsewhere [7]). In the case of arbitrary disturbances which are given functions of time, as a rule the differential equation corresponding to (15) will have to be solved, which in general will not be possible by elementary means.

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[6])   If an oscillator is to be "linearized" by giving the controlling mechanism a very large time constant, it is not permissible to introduce at the same time a constant oscillator voltage.

[7])   A treatise on oscillator noise, based upon the statistical fluctuations, has meanwhile been given by Aug. Blaquière, Comptes Rendus Ac. Sc. Paris **234**, 419-421, 710-712, 1140-1142, 1952 (Nos. 4, 7 and 11). Articles based on the Fourier spectrum of the noise source have been published by A. Spaelti, Bull. Schweiz. Elektrot. Vereins **39**, 419-428, 1948 and L. M. Lerner, Proc. Nat. Electronics Conference **7**, 275-287, 1951.

Summary. Starting from the supposition that the voltage produced by an oscillator is approximately sinusoidal and that the amplitude-controlling mechanism has no effect upon the phase of the oscillator voltage, the phase shift of that voltage due to an interfering pulse has been calculated. This is followed by an elementary explanation of the synchronization of an oscillator by a periodic disturbance differing in the duration of its cycle from that of the oscillator. By solving a differential equation the phase and frequency variations under the influence of a weak sinusoidal interference are calculated. As the strength of the interference increases, or the frequency difference decreases, the solution passes continuously into the state of synchronization. Further, the non-linearity of the phenomena is discussed, i.e. the dependency of the effect of an interference upon the presence of other interferences. This is followed by an explanation how the method followed can be applied for calculating the influence of noise voltages in the oscillator upon the frequency constancy of the oscillator.

PHOTOSYNTHESIS

by R. van der Veen.   581.132.1

In the final analysis, all life on this earth depends on the transformation of sunlight into chemical energy. This process, called photosynthesis, takes place exclusively in green plants and even there the output is small. If in this respect mankind could succeed in making itself independent of the vegetable world, the prospects for the energy provision of the whole world would be very favourably influenced. If we really intend to reach this goal it is important to obtain a fuller insight into the complicated process of photosynthesis; at present only the fundamental principles are known. It has been found that we can contribute to our knowledge by studying the fluorescence of living elements of the vegetable world.

The energy radiated from the sun to the earth amounts to about 10^{18} kWh a year. The energy consumption in a country as highly developed as the USA amounts only to about 10^{13} kWh a year, while the figure for the whole world is only 3 times as large as that for the USA. By far the largest part of the energy used by mankind has been derived since time immemorial from the coal and oil deposits found in the earth's crust, formed long ago by solar energy. These deposits however are not inexhaustible, and assuming that present-day technical development proceeds at the same speed, it has been calculated that after some centuries these deposits will have been completely consumed. (If we tried to replace the solar energy supplied to us every day by the energy stored in the oil and coal deposits we would consume the whole stock within 3 days!)

As might be expected science is already carrying out experiments to compensate the effects of this depletion. A modest step in this direction has been made by using atomic energy released from thorium and uranium but the available quantities of these elements are limited. If we could succeed in transmuting hydrogen into helium in a technically workable process, we certainly would postpone the day of exhaustion for many hundreds of thousands of years, but to achieve this we have still a long way to go.

It is obviously desirable to exploit the possibility of making direct use of solar energy. A small percentage of the energy in the world is generated in this way, viz. by making use of wind and water power, but it is improbable that the amount of energy so obtained can be considerably increased. Now, it is known that part of the total solar energy supplied is transformed by green plants into chemical energy. This process — photosynthesis — is the basis of all life and consists of the formation of carbohydrates from water and the carbonic acid of the atmosphere by the influence of sunlight. This photosynthesis results in the accumulation of energy to an amount of 10^{15} to 10^{16} kWh a year. However, this energy is lost again for the greater part without being used to an appreciable extent by mankind, by decomposition of the organic substances through bacteria, etc.

If we look at these figures it is evident that all our worries about the exhaustion of our energy deposits would disappear at once if we could only succeed in making a more efficient use of this photosynthesis. The first step in this direction is to obtain an insight into the very complicated system of reactions which constitute photosynthesis.

Photosynthesis

Photosynthesis is the most important chemical reaction taking place on earth. All life is dependent on the chemical energy stored in its own organic substance. The processes which govern life are kept going by the release of this energy by respiration, i.e. by the combustion of organic substance. Consequently a primary condition for the conservation of life is the formation of organic matter in one way or another. This formation takes place mainly by a single method, viz., by the influence of sunlight on a certain substance present in green plants — chlorophyll. (Certain organisms capable of using light for building organic substances are not green but close examination has shown that they do contain chorophyll, the green colour being suppressed by other colours.)

It is remarkable that the higher developed beings such as the animals, are unable to make direct use of photosynthesis. Not even in the laboratory has this proved possible. This means that all the higher developed creatures including mankind have to live parasitically on the green plants.

In general the chlorophyll present in the leaves of a plant is in the form of granules or chloroplasts, moving about in the plasma of the cell. In *fig. 1* an electron micrograph of some chloroplasts is shown. It is clearly to be seen that each
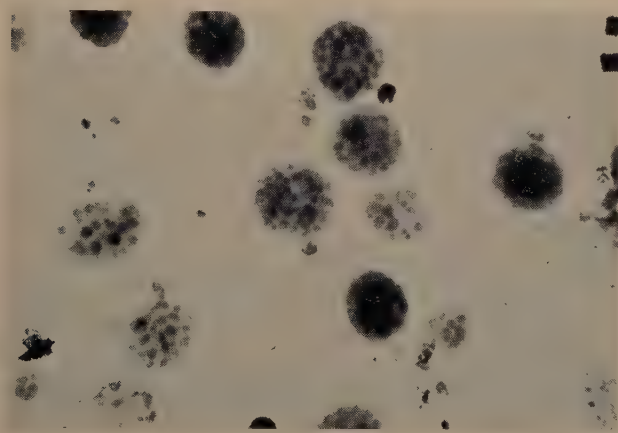

Fig. 1. Electron micrograph of chlorophyll granules (chloroplasts). The distance between the black squares in the right-hand top corner $= 1\ \mu$. It can be seen that the granules are built up from small corpusculae, the grana, embedded in the colourless stroma.

chloroplast consists of a number of lenticular bodies, the grana, embedded in a colourless gel, the stroma. *Fig. 2* shows a fragment of a chloroplast at a higher magnification.

The photosynthesis process as such takes place inside the chloroplasts. Schematically, photosyn-

thesis can be represented by the reaction:

$$CO_2 + H_2O \xrightarrow{\text{light}} (CH_2O) + O_2.$$

Though this formula looks very simple, the process that really happens is extremely complicated and

Fig. 2. Electron micrograph of a fragment of a chorophyll granule. Four grana are visible, embedded in a patch of stroma. The distance between the white squares in the right-hand bottom corner $= 1\ \mu$.

up till now it has been only partly unravelled. The final product, represented by (CH_2O), and which should be read as a collection of higher carbohydrates (sugars) is the organic substance which supplies the energy for the conservation of life.

Thus in fact photosynthesis consists of the reduction of CO_2 to cabohydrates, whereas the respiration process corresponds to the inverse reaction, i.e. the oxidation of carbohydrates to CO_2.

Mechanism of the reaction

It has been found that the chloroplasts alone cannot effect a photosynthesis of carbohydrates from CO_2 and water. After having been exposed to light, the chloroplasts do possess a certain reducing power, as can be concluded from the reduction of easily reducible substances as e.g. ferric salts to ferrous salts, during which process oxygen is liberated. In order to reduce carbon dioxide, however, a preliminary treatment of this substance is necessary. This preliminary treatment takes place in the protoplasm, independent of the light; for this reason it is called a dark-reaction. The light reaction proper, coupled with a number of dark-reactions, takes place in the chloroplasts.

From the foregoing it is already evident that the photosynthesis process as a whole is infinitely more complicated than is suggested by the simple equat-

ion given above. To clarify the picture — without undue complication — *fig. 3* illustrates how the reaction system looks according to present-day knowledge.

In many (largely unknown) intermediate stages the CO_2 from the atmosphere is converted into a

have succeeded in gaining a better insight into the various reactions which take place.

The first author has shown that under special circumstances it is possible to effect the CO_2 conversion by the chloroplasts, by adding diphosphopyridinenucleotide (DPN) to a suspension
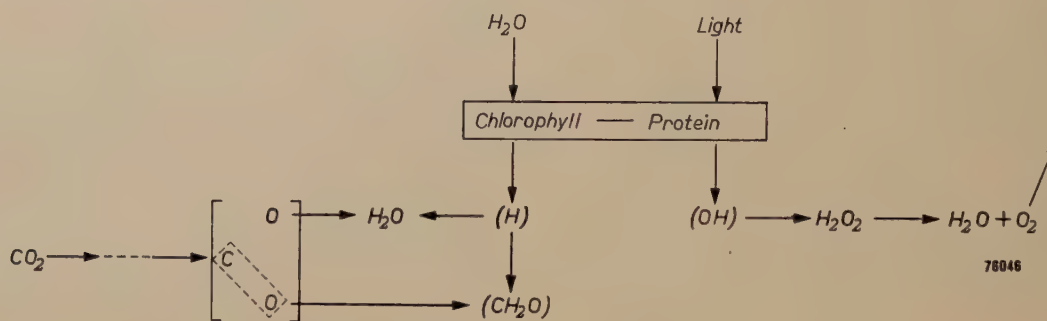
Fig. 3. Schematic reproduction of the photosynthesis process. The dark-reactions are on the left. On the right is the reduction of H_2O_2 to H_2O and O_2 by the enzyme catalase. The actual light reaction in the chloroplast narrows down to the decomposition of water into an (H)-component and an (OH)-component. All the oxygen which finally escapes originates from the water, not from absorbed CO_2.

complexity of unknown chemical compounds, indicated in the scheme of fig. 3 as $[C_O^O]$. At the same time water is decomposed by the chlorophyll-protein complex of the chloroplasts under the influence of light into a reducing component, indicated by the symbol (H) and an oxidizing component indicated by (OH). It has not yet been agreed as to what is meant exactly by these compounds; it is possible that we have to deal with H-atoms and OH-radicals.

The (H)-compound serves to reduce the $[C_O^O]$ complex, formed in the dark-reaction, to carbohydrates and water. The (OH)-components probably combine to H_2O_2-molecules, which, under the influence of the enzyme catalase (a substance, which has been isolated and is present in a relatively high concentration in every chloroplast) decompose immediately in water and oxygen. The oxygen escapes.

According to this scheme, the oxygen which finally escapes originates entirely from the water in the plant. From the symbolic equation at the beginning of this paper one might expect that half the liberated oxygen — or even more — originates from the CO_2, but experiments carried out with the oxygen isotope O^{18} have completely disproved this simple conception.

Much research work has been carried out to clarify those details of the photosynthesis which are still obscure. Among others, Ochoa[1]) and Calvin[2])

of chloroplasts. DPN is a complicated, phosphorus-containing organic substance, which seems to be present in all live cells and which is reducible. However, when other reducible substances and certain enzymes are present, it is easy to re-oxidize the reduced DPN. Ochoa showed that in the presence of reduced DPN and certain enzymes, CO_2 can combine with pyruvic acid to form malic acid. During this reaction the reduced DPN is oxidized, but it is immediately reduced again by the chloroplast which has been exposed to light. Thus, in a sense, the DPN acts as a catalyst. By the combined action of chloroplast, light energy, DPN and enzymes, the reaction always proceeds in the same direction, i.e. in the direction of the reduction of CO_2 to carbohydrates.

By making use of radio-active carbon, Calvin has proved, however, that the CO_2 fixed during the photosynthesis is first found, not in malic acid, but in the form of phosphoglyceric acid. After some time the radio-active C-atoms spread over a large number of other compounds. It may well be that these results are not opposed to those of Ochoa: in the sense that both results are based on the reduction of DPN by the chloroplasts. It is interesting to note that organic phosphates are necessary to effect photosynthesis, a fact which has been established by Wassink, Wintermans and Tjia[3]).

[1]) S. Ochoa, Symposia Soc. Exp. Biol. 5, 29, 1951.
[2]) M. Calvin, J. Chem. Educ. 26, 639, 1949.

[3]) E. C. Wassink, J. F. G. M. Wintermans and J. E. Tjia, Proc. Kon. Akad. Wet. Amst., Series C, 54, 496-502, 1951 (No. 5).

The question arises — why is the course of the photosynthesis reaction so extremely complicated? To answer this question it should be realized that in nature only those processes which lead to the final product in an efficient manner have a chance to succeed. In order to effect so radical a reaction as the synthesis of carbohydrates by means of the relatively small energy (light and heat) available, nature has to rely on subdividing the reaction into a very large number of intermediate reactions, the direction (sense) of which can be easily influenced by the circumstances. Now the presence of light and certain enzymes results in a slight shift of the equilibria in the plant cells towards the reducing direction. This requires only a very small energy-absorption per process. Moreover, it should be taken into account that the compounds formed not only play a part in the reactions considered, but also have very definite functions in the organisms, which results in the increase of the "efficiency" of the entire process.

Light saturation

The speed of photosynthesis is determined by the slowest reaction in the chain of the light and dark-reactions. The rate of progress of the light-reaction increases with the increase of the light intensity. In the temperature region where physiological processes can normally take place, i.e. between about 5—30 °C, this speed is almost independent of the temperature. The dark-reactions, on the contrary, are pure chemical reactions, of which the

rate of progress rapidly increases with the temperature. At a given temperature an increase of the light intensity will result in an increase of photosynthesis, but only as long as the light-reaction is the limiting factor. When one of the dark-reactions becomes limiting, an increase of the light intensity can no longer cause a change in the rate of photosynthesis: light saturation sets in. The higher the temperature, the greater the rate of progress of the dark-reactions, and light saturation will set in at higher intensities. This has been represented schematically in *fig. 4*.

Analysis of the light-reaction by means of fluorescence

As already demonstrated, only the fundamental principles of the process of photosynthesis are understood at present. The transitions which take place in the dark-reactions are still unknown and so is the light-reaction. But there is a possibility of becoming better acquainted with the latter.

Only part of the absorbed light energy is transformed in chemical energy during the light-reaction. In nature the output is only a few percent, whereas it has been possible to obtain an output of of about 30% or more in the laboratory when working with low light intensities. The remainder of the light energy is transformed into heat and for a small part into fluorescent radiation. The latter fact is most fortunate, for by studying the fluorescence, it is possible to follow the course of the light-reaction to a certain extent. It may be assumed that the relation between the energy given off as heat and the energy given off as fluorescence is always constant, whatever the percentage of the absorbed radiation energy used for the chemical transformation. Consequently the intensity of the fluorescence is a yardstick for the quantity of light energy bound chemically, and so for the output of the photosynthesis: the intensity of the fluorescence will be smaller according as the output is higher.

It is evident that the measurement of the intensity of the fluorescence can teach us only about the progress of the light-reaction. As far as the progress of the dark-reactions (including those following the light reaction) are concerned, this method will bring us no further. It may well be that part of the chemically bound energy is liberated in a later stage of the process as heat. This is the case e.g. if a leaf is exposed to light in the absence of CO_2. In that case there is no material that can be reduced and no oxygen is set free. However, the fluorescence is hardly any higher than it would be if CO_2 were
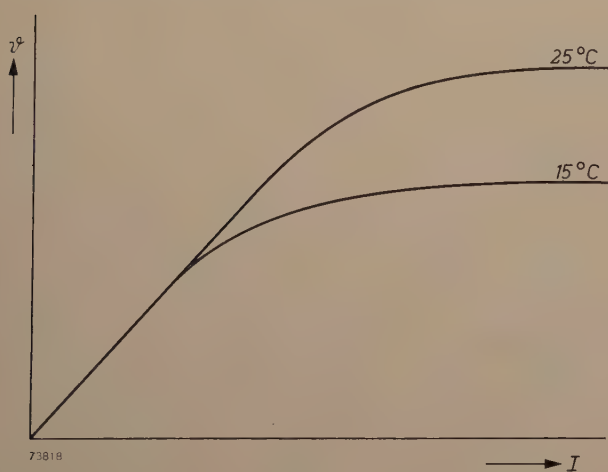

Fig. 4. Speed v of photosynthesis as a function of the illumination intensity I and the temperature. The light-reaction determines the speed in the sloping part of the curve, whereas the speed of the reaction is completely limited by the dark-reactions in the horizontal part, where light saturation sets in. The rate of progress of these dark-reactions increases as the temperature becomes higher. The light-reaction is almost independent of the temperature, the result being that light saturation at a higher temperature sets in only at a higher light intensity.

present. The fundamental light-reaction, i.e. the splitting-up of water into (H)- and (OH)-compounds, obviously takes place at the same rate, the presence or absence of CO_2 being immaterial. However, CO_2 being absent, these compounds combine again to water, during which process heat is generated and from a chemical point of view, it appears that nothing has happened.

The Philips Laboratory has carried out extensive researches on the phenomena which occur at the beginning of the illumination in the process of photosynthesis. In particular the CO_2-absorption and the fluoresence were studied [4]).
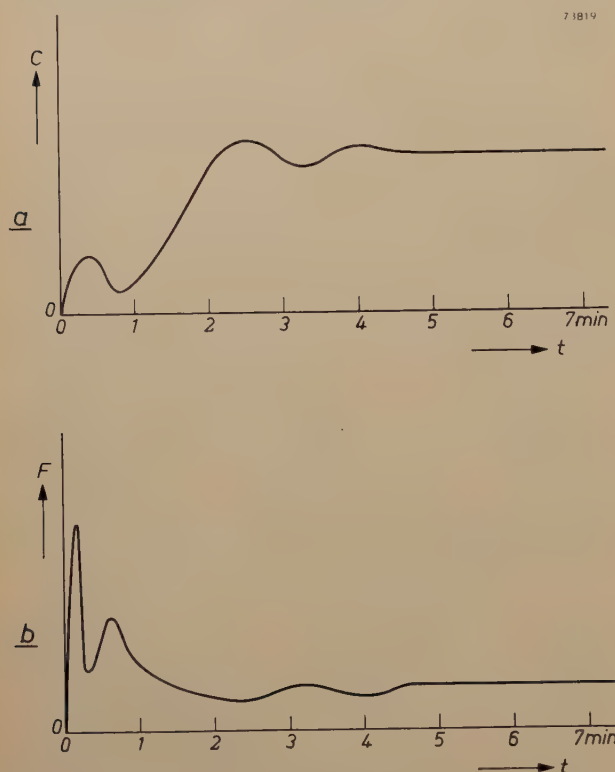
Fig. 5. a) Progress of the CO_2-absorption C per unit of time. b) Intensity of the fluorescence F, as function of the time t after applying the illumination. With the exception of the very early part, the two curves are nearly reflected images.

The general course of the CO_2-absorption is illustrated in *fig. 5a*. During the first moments (about 10 seconds) rather a large quantity of CO_2 is absorbed, but then the rate of absorption decreases quickly. About a minute after illumination is started, there is a gradual increase of the photosynthesis (adaptation period), until an almost constant level of the CO_2-absorption is reached, the height of the level being dependent on the intensity of illumination and the plant in question. The fluorescence shows, apart from the early beginning period, closely corresponding variations

[4]) See, for example, R. van der Veen, Physiologia Plantarum **4**, 486-494 1951 (No. 3).

(fig. 5b): the CO_2 absorption at the beginning is reflected in a short-period decrease, and the adaptation period as a continued gradual decrease. Fluctuations in the photosynthesis which may set
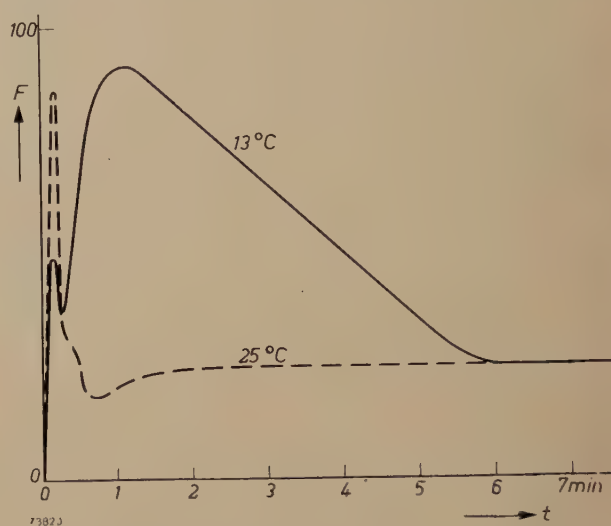

Fig. 6. Intensity of fluorescence F as function of the time t after applying the illumination, at two temperatures, 13 °C and 25 °C. Though the ultimate level is the same in both cases, it is attained much earlier at the higher temperature. The measurements have been carried out on a dahlia-leaf.

in later are completely reproduced in the intensity of the fluorescence. In the very early part, however, the fluorescence shows a sharp peak, the relation of which to the photosynthesis is still obscure.

The influence of the temperature, the light intensity and the duration of the dark period preceding the illumination has been investigated. In the early stages the rate of photosynthesis is much lower at low temperatures (say, 13 C°) than at high
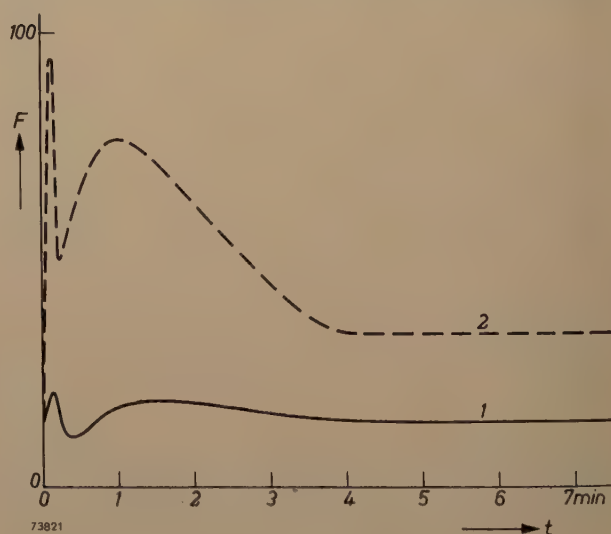

Fig. 7. Intensity of fluorescence F as function of the time t after applying the illumination at two intensities, differing by a factor 3. Curve *1* appertains to the low intensity, curve *2* to the high one. Measurements have been carried out on a tomato leaf.

temperatures (25 °C). This is reflected in the high value of the fluorescence at 13 °C as compared with that at 25 °C, during the first minutes *(fig. 6)*. It can be seen, however, that the ultimate level of the fluorescence is not influenced by the temperature. This confirms the assertion that the light reaction depends little on the temperature.

The intensity of the fluorescence increases when the light intensity is increased, as would be expected. It is remarkable, however, that the increase is much more marked when the illumination is first begun than in later periods *(fig. 7)*.
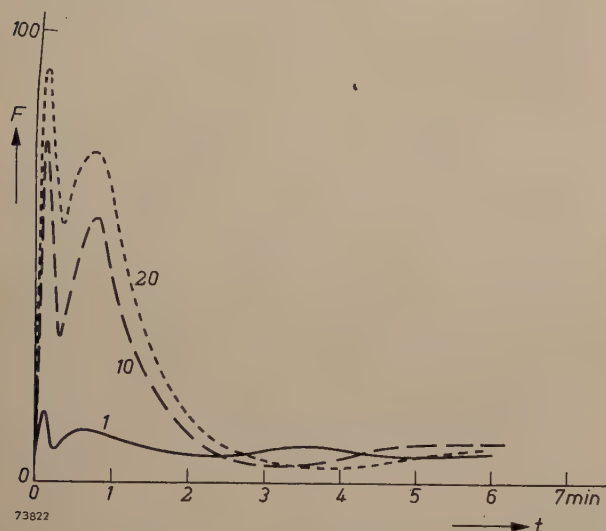

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Fig. 8. The influence of the length of the preceding dark period on the fluorescence curve. The figures against each curve represent the length of the period in minutes. The measurements were carried out on a tomato leaf.

The influence of the preceding dark period is represented in *fig. 8*. The longer the period, the higher are the various maxima. This agrees completely with what has been found by other, more direct means. It seems that the plant remembers the previous illumination, provided this took place not longer than 20 minutes previously.

It is impossible as yet to have a clear picture of the physical and chemical background of all these phenomena, but it is certain that the study of fluorescence is of material aid to investigatioms into photosynthesis.

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Summary. Photosynthesis, the basis of all life on earth, takes place only in green plant parts. The substance called chlorophyll is an essential link in the transformation process of light into chemical energy. The process of photosynthesis, essentially consisting of the reduction of CO_2 to carbohydrates and accompanied by the liberation of oxygen—and as such to be considered as the inverse of the respiration process — is very complicated. The actual light-reaction takes place in the chloroplasts of the plant, but it is preceded by a large series of dark-reactions, occurring partly in the protoplasm, which effect a preliminary transition of the CO_2 into an easily reducible substance. The light-reaction can be considered as the splitting up of water, under the influence of light, into a reducing and an oxidizing compound; the former effecting the reduction of CO_2 to carbohydrates and water, and the latter ultimately liberating oxygen. It has been found that the presence of organic phosphates in the dark-reactions is very important. The output of the photosynthesis in nature is only a small percentage, whereas in the laboratory, at low light intensities, an output of 30% and higher has been attained. A large part of the absorbed light energy is lost as heat, and a small part as fluorescent radiation. By studying the fluorescence of the green parts of plants, science has succeeded in investigating the early part of the light-reaction. The results of the fluorescence research indicate that the intensity of fluorescence is often a measure of the process of photosynthesis.

AN OSCILLATOR WITH CONSTANT OUTPUT VOLTAGE

by L. ENSING and H. J. J. van EYNDHOVEN.   621.396.615:621.316.722.1

Using the general principle of a regulating system, a valve oscillator can be made to work in such a way that the output voltage remains substantially constant. An application regarding an installation for the calibration of valve voltmeters is described.

For the generation of alternating voltages with fixed orc ontrollable frequency, use is often made of oscillators based on an electronic valve (triode or pentode), an oscillatory circuit (*LC* circuit) and a feedback element. An oscillation generated in this way can, however, show fluctuations in amplitude and in frequency and deviations from the sinusoidal waveform which are most undesirable, especially for measuring and calibrating purposes.

We shall describe below a method which is primarily directed towards making the amplitude of the oscillator independent of disturbing influences such as changes in the supply voltages of the valve, in the mutual conductance of the valve and in the impedance of the *LC* circuit. With this method the valve is working under conditions which are also favourable for the frequency to remain constant and for the generation of a purely sinusoidal voltage.

Regulating principle applied to an oscillator

A method on which many regulating systems are based [1] consists in the comparison of the regulated quantity x with a similar but constant reference quantity a_{ref}, in that the difference $a_{ref}-x$ is made to control the apparatus in such a way that $|a_{ref} - x|$ is reduced.

If the quantity which has to be kept constant and the reference quantity are not similar, then an auxiliary quantity will have to be derived from the first mentioned one which is similar to a_{ref}. This is the case in the matter under discussion: the alternating output voltage of a valve oscillator has to be kept constant in amplitude (V_0), but as reference voltage we can of course only take a direct voltage (E_{ref}). By rectification a direct voltage E_0 which is proportional to V_0 can be derived from the oscillator voltage and the difference $E_{ref} - E_0$ can function as control voltage, i.e. it can control the oscillator in such a way that the difference $|E_{ref} - E_0|$, irrespective of its cause, becomes smaller. To attain this, the control voltage can work, for example, as a bias in the control-grid circuit with such polarity

that the working point on the valve characteristic is shifted to a region of greater or smaller mutual conductance accordingly as the voltage amplitude V_0 was either too small or too large.

Influence of the valve adjustment

Before elaborating on this thought it should be recognized that the limitation of the oscillation amplitude is effected by the working of the regulating system and not — as in the case of the usual *LC* oscillators with a grid capacitor and grid leak — by the occurrence of grid current. The latter biases the control grid to such an extent that the valve is cut off during part of the cycle (class C operation) and the effective mutual conductance decreases. The anode current thus becomes pulse-shaped and consequently contains strong harmonics; these harmonics also occur in the circuit voltage, although to a lesser degree.

By using a regulating system, however, one can ensure that the valve remains operating in class A; thus operation is limited to a small, substantially straight part of the i_a-v_g characteristic, resulting in much less distortion of the circuit voltage.

The last is not only important if a sinusoidal output voltage is desired, but also because distortion is one of the causes of frequency fluctuations. As Horton has shown [2], distortion results in a phase shift in the voltage which is fed back to the grid, so that the frequency f for which the oscillatory condition is complied with, is no longer the resonant frequency f_0 of the *LC* circuit. The difference between f and f_0 increases as the distortion becomes larger and will therefore not be constant if the supply voltage fluctuates, if the valve heats up, and if it ages in the long run.

Another cause of the frequency shift may lie in the well-known fact that the inter-electrode capacitances of a valve — especially the grid-to-cathode capacitance — depend on the space charge. In the circuit to be discussed the regulating system will keep the direct current which flows through the

[1] See e.g. J. H. Roosdorp, On the regulation of industrial processes, Philips tech. Rev. Vol. 12, 221-227, 1951.

[2] J. W. Horton, Vacuum tube oscillators. A graphical method of analysis, Bell. Syst. tech. J. 3, 509-524, 1924.

valve approximately constant in case of mains voltage fluctuations and when the valve ages, so that the influence of this source of frequency fluctuations is reduced.

More detailed description of the regulating circuit

In various ways it is possible to effect the limitation of the amplitude, so that the oscillator remains operating in class A. In one well-known method a small incandescent lamp [3]) (or another device having a resistance dependent on temperature) is used in one arm of a Wheatstone bridge; this bridge can consequently only be in equilibrium at one particular value of the input voltage.

Another method has been stated by Arguimbau [4]), who uses a separate rectifier which converts an alternating voltage proportional to the oscillator voltage into a direct voltage, the latter serving as bias in the control grid circuit.

Taking the discussion of the regulating principle given above as a starting point, we can describe our method [5]) as one in which two currents are compared instead of two voltages. Like Arguimbau we use an auxiliary rectifier; it consists of a diode D_1 and a blocking capacitor C_a (fig. 1a),

connected in series between the cathode and the anode of the oscillating pentode P. Consequently a pulsating voltage v_d is produced across the diode, with the anode negative with respect to the cathode (fig. 1b); the amplitude of v_d is $2 V_0$ and its average value $E_0 = -V_0$. A capacitor C_g is connected between the points G and K (fig. 1a) of the grid circuit; G is connected to the anode of the diode via a resistor R_1 and to a point having a positive voltage, the reference voltage E_{ref}, via a resistor R_2. Currents I_1 and I_2 flow through these resistors and with a negligible grid current the following formula applies:

$$ I_1 + I_2 = 0. \quad \cdots \cdots \quad (1) $$

If E_{g1} is the direct voltage across the capacitor C_g, then $I_1 = (E_0 - E_{g1})/R_1 = -(V_0 + E_{g1})/R_1$ and $I_2 = (E_{ref} - E_{g1})/R_2$. Substitution of these values in (1) gives:

$$ V_0 = \frac{R_1}{R_2} E_{ref} - \frac{R_1 + R_2}{R_2} E_{g1}. \quad \cdots \quad (2) $$

E_{g1} — the control grid bias required to make the valve oscillate with the amplitude V_0, at given supply voltages and circuit impedance — depends on the valve characteristics. It is, however, possible to make the first term in the right-hand member of (2), in which E_{g1} is absent, preponderate strongly with regard to the second term, for example by choosing E_{ref} high with respect to E_{g1}; in first approximation we then find that

$$ V_0 = \frac{R_1}{R_2} E_{ref} , $$

viz. in first approximation the amplitude of the output voltage depends solely on the ratio of two fixed resistances and on one direct voltage, and consequently not on the valve characteristics and the circuit impedance. For a more accurate analysis, however, it is necessary to take into account also the influence of the last-mentioned quantities.

The stabilisation factor

The direct anode voltage, the screen-grid voltage and the heater voltage applied to the pentode, and also the reference voltage E_{ref}, are as a rule obtained from the A.C. mains and therefore show the same relative fluctuations as the mains voltage, unless special measures have been taken to prevent this.

It is obvious that, if V_0 is to be kept constant, E_{ref} should in any case be carefully stabilised, for example by means of one or more voltage reference tubes. Fig. 1a shows that the screen-grid voltage
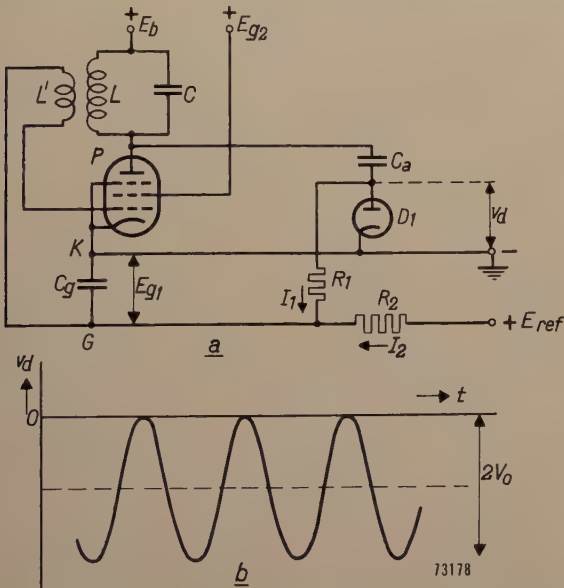
Fig. 1. a) Circuit for automatically stabilising the output voltage of an oscillator. P pentode. LC oscillatory circuit. L' feedback coil. C_a blocking capacitor and D_1 diode, rectifying the alternating anode voltage to a pulsating voltage v_d (represented in b as a function of the time t). E_{ref} reference voltage. C_g grid capacitor.

[3]) L. A. Meacham, The bridge-stabilized oscillator, Proc. Inst. Rad. Engrs. **26**, 1278-1294, 1938.
[4]) L. B. Arguimbau, An oscillator having a linear operating characteristic, Proc. Inst. Rad. Engrs. **21**, 14-28, 1933.
[5]) This corresponds to a large extent with the triplex diode circuit used in some radio receivers for delayed automatic gain control; see e.g. J. Deketh, Fundamentals of radio-valve technique I, pp. 329-332 (Philips Technical Library, 1949).

can be obtained from the same voltage source without any difficulty; it is then no longer necessary to be concerned regarding the influence of fluctuations on the screen-grid voltage.

The exact value of the direct voltage in the anode circuit is not a factor of decisive influence; the properties of a pentode are sufficiently independent of this voltage. Heater voltage fluctuations, however, result in variations in the mutual conductance of the valve characteristic; they can be considered together with changes in mutual conductance due to other causes.

Finally, we have yet to consider the influence of changes in the impedance in the anode circuit.

If the mutual conductance S changes by an amount ΔS (due to a change in the heater current or by ageing of the valve), the original value is restored by changing E_{g1} by an amount

$$\Delta E_{g1} = -\frac{\partial E_{g1}}{\partial S} \Delta S,$$

in which $\partial E_{g1}/\partial S$ can be determined from the $S = f(E_{g1})$ characteristic. From equation (2) follows:

$$\Delta V_0 = -\frac{R_1 + R_2}{R_2} \Delta E_{g1} = \frac{R_1 + R_2}{R_2} \frac{\partial E_{g1}}{\partial S} \Delta S.$$

We now find for the ratio of the relative change of the mutual conductance to that of the amplitude, i.e. the stabilisation factor for mutual conductance changes:

$$a_S = \frac{\Delta S/S}{\Delta V_0/V_0} = \frac{V_0}{S} \frac{R_2}{R_1 + R_2} \frac{\partial S}{\partial E_{g1}} . \quad . \quad (3)$$

The alternating voltage (amplitude V_0) across the impedance Z in the anode circuit is produced by the alternating anode current (amplitude $I_a = V_0/Z$), which in its turn is caused by an alternating voltage (amplitude V_{g1}) on the control grid: $I_a = SV_{g1}$. It will be assumed that this alternating grid voltage, which is due to the feedback, is a fraction t of the alternating anode voltage: $V_{g1} = tV_0$. To maintain the oscillatory state, the known oscillatory condition should be complied with:

$$SZt = -1, \quad . \quad . \quad . \quad . \quad . \quad . \quad (4)$$

which expresses that the total amplification for one round trip through the oscillator should be exactly equal to unity.

The impedance Z is variable in oscillators with adjustable frequency; moreover it varies with changes in the load on the oscillator and it can also vary due to incidental causes, for example by increase of the losses caused by dust and moisture. We see from (4) that changes in Z with constant t

are compensated by equal relative changes in S. Consequently the stabilisation factor for impedance variations,

$$a_Z = \frac{\Delta Z/Z}{\Delta V_0/V_0},$$

is equal to the value of a_S found in (3).

For the sake of completeness we shall finally consider the case of a non-stabilised screen-grid voltage.

It follows from (2) that a change ΔE_{g1} of the control-grid bias results in a change of the amplitude V_0 by an amount ΔV_0 of the following magnitude:

$$\Delta V_0 = -\frac{R_1 + R_2}{R_2} \cdot \Delta E_{g1}.$$

For a pentode it holds that with constant anode current a change ΔE_{g2} in the screen-grid voltage is compensated by a change $\Delta E_{g1} = -\Delta E_{g2}/\mu_{g2g1}$ in the control-grid voltage; μ_{g2g1} is a factor larger than unity (e.g. 18 with a type EBF 80 pentode).

A change in the screen-grid voltage might be thought to influence the mutual conductance. However, this is barely so with a pentode; the characteristics of the anode current as a function of E_{g1}, with E_{g2} as parameter, run almost parallel and consequently the mutual conductance which is required for oscillating is reached at a certain value of the anode current which is substantially independent of E_{g2}.

Consequently $\Delta E_{g2}/\Delta V_0$ (the regulating factor) becomes

$$\frac{\Delta E_{g2}}{\Delta V_0} = \frac{R_2}{R_1 + R_9} \mu_{g2g1} .$$

For the ratio of the relative changes, the stabilisation factor for changes in screen-grid voltage, we find:

$$a_{Eg2} = \frac{\Delta E_{g2}/E_{g2}}{\Delta V_0/V_0} = \frac{V_0}{E_{g2}} \cdot \frac{R_2}{R_1 + R_2} \mu_{g1g2}.$$

Numerical example

In agreement with a circuit we will discuss presently, we have selected as a valve the EBF 80 pentode. Let V_0 be (say) 40 V and $Z = 30$ kΩ, then I_a is 1,33 mA. With $t = {}^1/_{20}$, V_{g1} has to be 2 V and the required mutual conductance is then $S = 1·33/2$ mA/V = 0.67 mA/V.

Keeping the screen-grid voltage constant at a value of 85 V, the mutual conductance of 0.67 mA/V is obtained, according to the valve data, at a grid bias $E_{g1} = -4$ V, and amounts to $\partial S/\partial E_{g1} = 0.6$ mA/V² at this adjustment.

Choosing also 85 V for the reference voltage, and $R_1 = 1$ MΩ, then from (2) a value of 2.5 MΩ follows for R_2.

According to (3) the stabilisation factor for mutual conductance and impedance changes now becomes:

$$|a_S| = |a_Z| = \frac{40}{0.67 \times 10^{-3}} \cdot \frac{2.5}{1 + 2.5} \cdot 0.6 \times 10^{-3} = 26.$$

In this example the relative variations in the alternating voltage amplitude are consequently $1/26$ of the relative mutual conductance or impedance variations. This important improvement is obtained by means of only a few extra circuit elements. Considerably larger values of the stabilisation factor can be obtained by connecting an amplifier between the oscillatory circuit and the regulating device, as will be shown presently.

It should be observed that although stabilisation may be obtained to a certain degree with a conventional LC oscillator by stabilising the supply voltages (including the heater voltage), changes in mutual conductance by the ageing of the valve and impedance variations in the circuit may still occur. With the regulating circuit described above these influences are also kept in check.

Moreover, the regulating system keeps the valve constantly operating in class A. This not only offers the advantage that the distortion of the circuit voltage is very small — an advantage already mentioned above — but also that the interelectrode capacitances of the valve remain constant when the mains voltage fluctuates or the tube ages, since the anode current is kept constant in that case.

Adjustment and modulation of the alternating voltage amplitude

By making the reference voltage E_{ref} controllable, a simple means is obtained of varying V_0 or adjusting it to a given value. The graph showing V_0 as a function of E_{ref} in accordance with equation (2) is given in *fig. 2*.
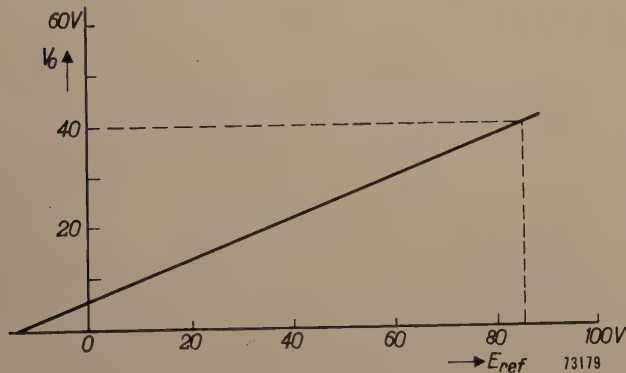
Fig. 2. Graphic representation of equation (2): amplitude V_0 of the alternating voltage versus reference voltage E_{ref} for $R_1 = 1$ MΩ, $R_2 = 2.5$ MΩ and $E_{g1} = -4$ V.

If E_{ref} is varied around a certain value, the amplitude of the output voltage is modulated. Use can be made of this if a modulated output voltage is required. Fig. 2 also represents the modulation characteristic, provided no grid current is flowing. If grid current does flow, equation (1) is no longer

applicable; the modulation characteristic then becomes less steep and consequently shows a sharp bend (*fig. 3*). Therefore the fact whether grid current flows or not, determines the greatest modulation depth at which no distortion occurs.
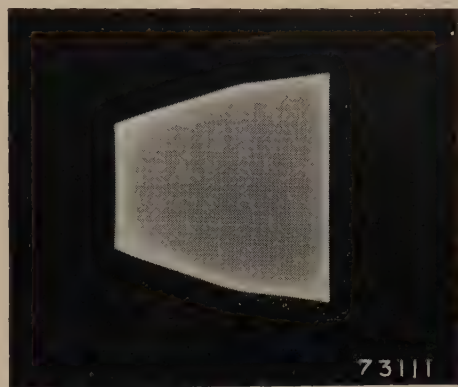
Fig. 3. Oscillogram of the alternating voltage of an oscillator in accordance with fig. 1a, which is being modulated with a sinusoidal voltage. The voltage for the horizontal deflection is likewise sinusoidal and in phase with the modulating voltage. For registering this oscillogram, the valve was so adjusted that grid current occurred, which caused a sharp bend in the envelope.

Plotting the modulation characteristic (or displaying the output voltage on an oscilloscope) is the simplest method of knowing whether grid current flows or not. The more usual method, i.e. to connect a voltmeter parallel to the grid resistor, would increase the stray capacitances to such an extent that the working point of the valve would be shifted considerably in order to remain in compliance with the oscillatory condition, so that the measurement would then be taken under very different conditions.

An upper limit is also set to the modulation frequency viz. by the tendency towards squegging.

Squegging and starting

The operation of the oscillator is determined mainly by the magnitudes of the capacitances C_g and C_a (fig. 1a). The capacitor C_g in the grid circuit, for example, together with the capicitance between control grid and earth, forms a voltage divider; if C_g is too small, a large proportion of the voltage induced in the feedback coil is lost in C_g. On the other hand, C_g should not be too large either — and this also applies to C_a — since the control voltage is then built up so slowly that there is a risk of squegging; the normal oscillating process is then periodically interrupted at a frequency f_{rel}, depending on a relaxation time.

It is possible to investigate whether there is a risk of squegging by modulating the oscillator in the above-mentioned way with an alternating voltage of constant amplitude and variable fre-

quency and by displaying the modulated output voltage on an oscilloscope. If the modulation frequency approaches f_{rel}, a deepening of the modulation is observed. This puts an upper limit to the modulation frequency. A square-wave modulating voltage can also be used; the tendency towards squegging can in that case be judged from the damping of the oscillations present in the envelope of the modulated output voltage wave (fig. 4).
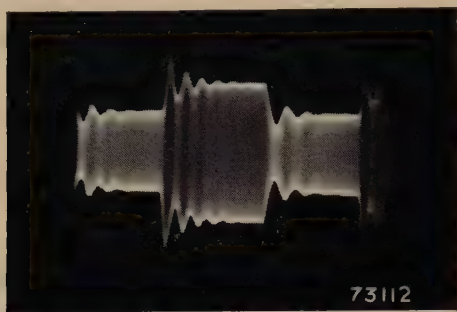

Fig. 4. Oscillogram of the voltage across the LC circuit of an oscillator in accordance with fig. 1a, modulated with a square-wave voltage. The tendency towards squegging can be judged from the damping of the oscillations present in the envelope.

If C_g and C_a have been so chosen that the oscillator (connected in accordance with fig. 1a) can

the resistor R_1 and the diode D_1, while the greater part flows through the feedback coil and further as grid current through the pentode. Due to the last fact the path between the control grid and cathode has a differential resistance which is much smaller than the impedance of C_g; consequently only a small part of the growing alternating voltage which is induced in the feedback coil reaches the control grid; moreover it does not have the correct phase. For this reason the oscillating process cannot start.

To remove this obstacle, provision should be made that I_2 cannot flow as grid current. To attain this, a diode D_2 can be incorporated (fig. 5a), offering an easy path to I_2, and a resistor R_3, increasing the resistance in the path via the control grid.

Fig. 5b shows that a single valve, containing a pentode system and two diode systems, such as the above-mentioned EBF 80, can be used to advantage. The reference voltage and the screen-grid voltage are stabilised at 85 V by one 85 A1 voltage reference tube [6]. Consideration is being given to the application of a system similar to that just mentioned in various measuring apparatus brought on the market by Philips.
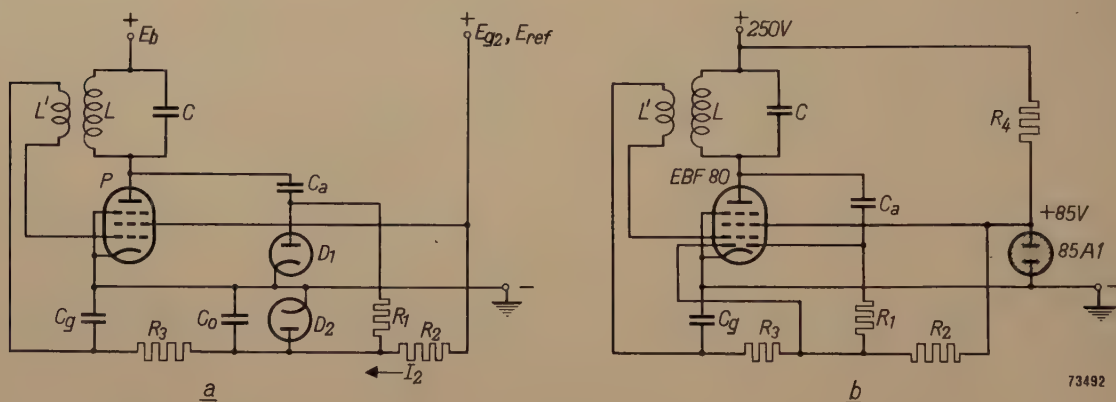

Fig. 5. a) As fig. 1a, but with the addition of a diode D_2 and a resistor R_3, which prevent the current I_2 from flowing across the control grid immediately after switching on; this would hamper the oscillator from starting properly. C_0 smoothing capacitor. Meaning of the other letter references as in fig. 1a. b) Practical example of the principle with an EBF 80 valve, in which the pentode P and the diodes D_1 and D_2 are combined. The reference voltage, also screen-grid voltage, is kept constant by a 85A1 voltage reference tube, with the series resistor R_4.

oscillate without tendency towards squegging, it apparently has some difficulty in starting.

The cause of this is the following: Immediately after switching on, the valve by no means oscillates at full strength and the alternating anode current is still negligible; the current I_2, due to the reference voltage source, now flows for a small part through

Installation for the calibration of valve voltmeters

The stabilisation of an oscillator voltage has already found practical application in an installation for the calibration of valve voltmeters.

Fig. 6 shows this installation for the calibration

[6] T. Jurriaanse, A voltage stabilizing tube for very constant voltage, Philips tech. Rev. 8, 272-277, 1946.
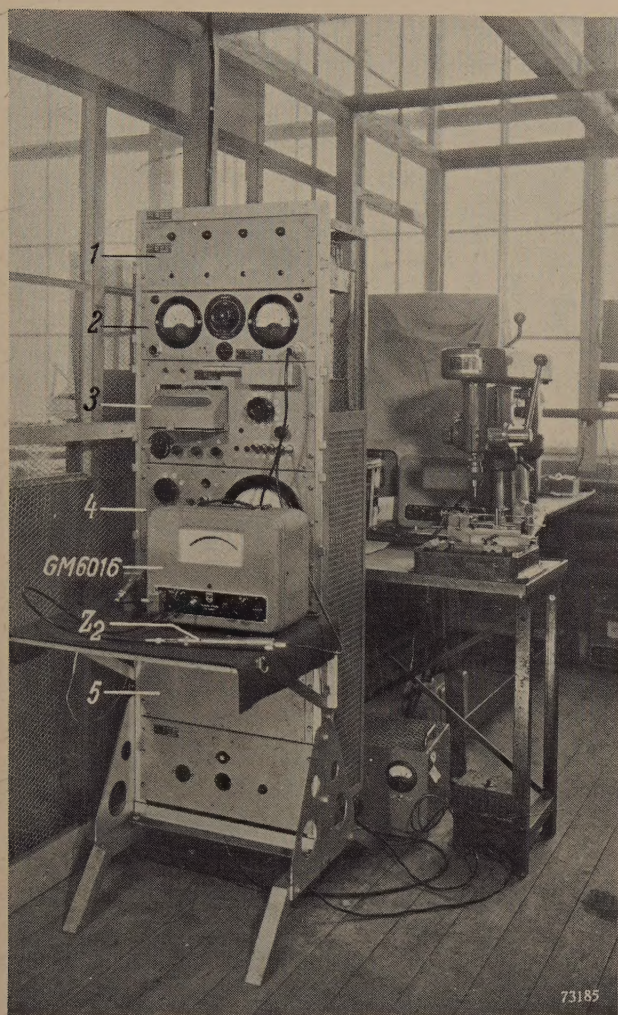
Fig. 6. Calibrating rack for the electronic millivoltmeter type GM 6016. *1* panel with four stabilised supply rectifiers. *2* panel for the adjustment of the supply voltage (50 c/s) for the meter to be calibrated. *3* panel with calibration voltages from 3 mV to 300 V, 100 kc/s, for the calibration of meter GM 6016 and the accessory attenuator (Z_2). *4* panel for checking the frequency characteristic. *5* panel with twelve oscillators (frequencies 1 kc/s to 30 Mc/s), with automatically stabilised output voltage. To the right, the drilling machine for drilling the catch holes in the calibrated attenuator.

of millivoltmeters type GM 6016, an instrument suitable for frequencies ranging from 1000 c/s to 30 Mc/s. The input voltage required to fully deflect the meter amounts to 3 mV at maximum sensitivity [7]; this voltage can be increased in steps up to 1000 V by means of a capacitive attenuator.

A similar installation is in use for the calibration of millivoltmeter type GM 6005; this is a meter suitable for frequencies from 20 c/s to 1 Mc/s which gives full deflection at voltages from 10 mV to 300 V. In the following we shall only discuss the calibration installation for type GM 6016.

[7] It is only with regard to sensitivity that type GM 6016 differs from type GM 6006 described before, which already gives full deflection at 1 mV (H. J. Lindenhovius, G. Arbelet and J. C. van der Breggen, Philips tech. Rev. **11**, 206-214 1950).

Two checks can be carried out with the installation under discussion:
1) absolute calibration of the meter including the attenuator, and
2) checking of the frequency characteristic.

An oscillator O_0 for 100 kc/s (*fig. 7*) is mounted in panel 3 (fig. 6) for the first-mentioned calibration; it is connected to an amplifier A_0, supplying a voltage with an r.m.s. value of 300 V. This voltage is stabilised by a regulating device R as discussed above; the components C_a, D_1, C_g, R_1 and R_2 correspond with those correspondingly marked in fig. 1a. The only difference between the two circuits is the presence of an amplifier in the one under discussion, which makes no difference to the principle, but as observed before, improves the stabilisation factor. The reference voltage E_{ref} is obtained from a stabilised direct voltage source of 310 V via a potentiometer.

Panel 5 (fig. 6) contains twelve oscillators, for the frequencies 1, 5, 100 and 1000 kc/s and 1, 3, 10, 15, 18, 20, 25 and 30 Mc/s, for checking the frequency characteristic [8]. The circuit voltage of these oscillators is stabilized at 30 V (r.m.s. value), by application of the regulating principle discussed above.
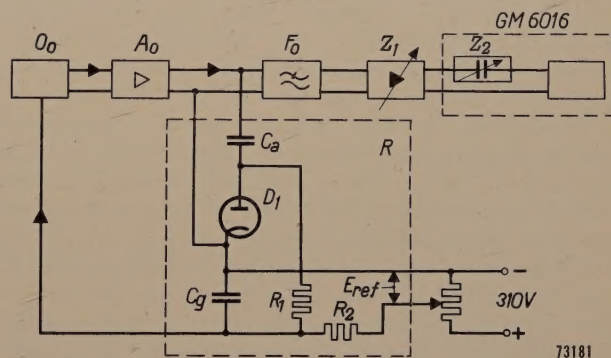

Fig. 7. Circuit diagram for the calibration of the meter GM 6016 and its attenuator Z_2. O_0 oscillator with frequency 100 kc/s. A_0 amplifier. R, with the component parts C_g, D_1, C_a, R_1 and R_2: unit for the producing of a control voltage (see fig. 1a) which stabilizes the output voltage of A_0. The reference voltage E_{ref} is adjustable. F_0 low-pass filter. Z_1 attenuator adjustable in calibrated steps.

Absolute calibration and checking of the attenuator at 100 kc/s

Although it actually falls outside the scope of this article, a brief outline of the calibration procedure is given below.

An attenuator Z_1 (fig. 7) belonging to the calibrating installation and to which 300 V at 100 kc/s is applied, delivers calibrating voltages of 300, 100, 30 and 10 V to a series of sockets, and in addition to this, by means of a switch, voltages of 3 and 1 V and 300, 100, 30, 10 and 3 mV can be obtained.

[8] The presence of two oscillators for 1000 kc/s (= 1Mc/s) will be explained later.
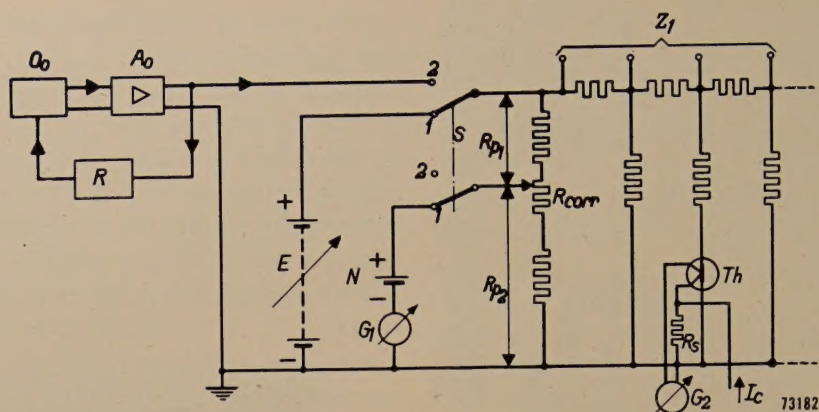
Fig. 8. Device for adjusting the 100 kc/s input voltage of attenuator Z_1 exactly to 300 V.
   Switch S in position 1: After the correction potentiometer R_{corr} has been set in accordance with the temperature of the standard cell N_1, the direct voltage E is so adjusted that the deflection of the galvanometer G_1 becomes zero, E is then exactly 300 V. The deflection of the galvanometer G_2, which is connected to four thermocouples Th connected in series (of which only one is shown in the diagram), is reduced to zero by means of the current I_c, flowing through the resistor R_s.
   Switch S in position 2: The reference voltage E_{ref} of the oscillator-amplifier O_0-A_0-R (see fig. 7) is so adjusted that G_2 again returns to zero at exactly the same value of I_c as before. The r.m.s. value of the voltage at the input of Z_1 is then exactly 300 V.

First the calibrating voltage of 3 mV is applied direct to the terminals of the meter GM 6016 which has to be checked; the meter is so corrected that it gives the proper deflection. Next one changes over to the higher calibrating voltages, the capacitive attenuator belonging to the meter (Z_2, fig. 7) being interconnected. This attenuator, which is continuously variable, is each time adjusted in such a way that the meter gives the correct indication; at these positions holes are drilled, so that a catching device, consisting of a steel ball, which is pressed into the hole by a spring, stops the attenuator.

The question, however, arises how it can be ascertained that the high-frequency alternating voltage at the input of Z_1 is exactly 300 V. This is checked by means of a direct voltage, by comparison with the voltage of a standard cell. The procedure is as follows.

The input side of Z_1 is disconnected from the amplifier A_0 and connected to an adjustable direct voltage source $E \approx$ 300 V (fig. 8). To make E exactly 300 V it would of course be possible to use a moving-coil voltmeter, but greater accuracy can be obtained by comparing a given fraction of E with the voltage of a standard cell in a compensation circuit. Therefore, the direct voltage source is shunted by a potentiometer, consisting of the (substantially) fixed resistors R_{p1} and R_{p2} (fig. 8). The ratio of these resistances has been so selected that the voltage across R_{p2} is equal to that of the standard cell N_1 if E is exactly 300 V. To make $E = 300$ V, E is varied until the deflection of the galvanometer G_1 has become zero.

Since the voltage of a standard cell depends to a certain degree on the temperature, the ratio $R_{p1} : R_{p2}$ has to be variable to a slight extent; the small correction potentiometer R_{corr} serves this purpose. This potentiometer is provided with a temperature scale; it is set to the temperature read on a thermometer inserted in the box of the standard cell (Philips type GM 4569, fig. 9).

If a direct voltage of exactly 300 V at the input of the attenuator Z_1 has thus been obtained, a certain current will flow through each of the resistors of Z_1. One of these currents is measured in relative value by means of four thermocouples Th connected in series (four to increase the sensitivity)
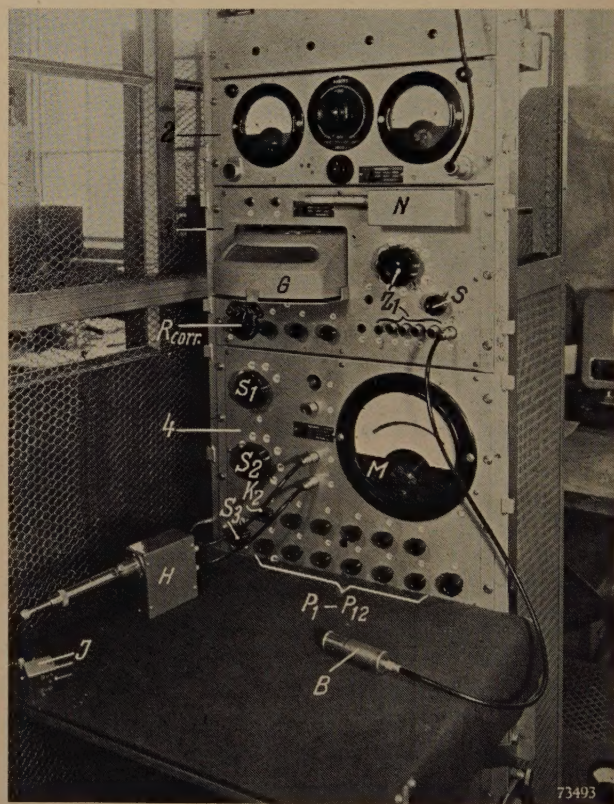

Fig. 9. Close-up view of the panels 2, 3 and 4.
   Cf. fig. 8.: N standard cell (type GM 4569) with thermometer. G galvanometer serving either as G_1 or as G_2. S switch changing over from direct voltage to alternating voltage. Z_1 attenuator adjustable in steps. B socket to which the attenuator of the meter GM 6016 is plugged.
   Cf. fig. 10: S_1, S_2, S_3 frequency selector switches. P_1-P_{12} potentiometers for adjusting the oscillator voltages to the correct value. K_2 coaxial cable. H probe with germanium diode and capacitive voltage divider (Ge_2 and C_1-C_2 in fig. 10). The meter GM 6016 to be checked is connected to the plug J. M meter for the comparing the oscillator voltages.

and a galvanometer G_2 (fig. 8). This too, is effected by a zero method, for the galvanometer circuit includes a resistor R_s which is traversed by an adjustable auxiliary direct current I_c. This current is adjusted in such a way that the deflection of the galvanometer G_2 becomes zero, i.e. that the voltage across R_s exactly compensates the output voltage of the thermocouples; I_c then is a measure of the current flowing through the heaters of the thermocouples. Now Z_1 is switched over from the direct voltage source to the amplifier A_0, which supplies an alternating voltage of 100 kc/s, and the reference voltage E_{ref} is adjusted until the deflection of the galvanometer G_2 again becomes zero. In that case the alternating voltage at the input of Z_1 has an r.m.s. value of exactly 300 V (0.1% difference already makes G_2 deflect 15 gradations), and at the output terminals of Z_1 the calibrating voltages 300 V, 100 V, 30V, ...3 mV are available. The attenuator Z_1 has been corrected for stray capacitances, so that the attenuation is the same with an alternating voltage of 100 kc/s as with a direct voltage.

The deflection of the meter GM 6016 depends on the form factor of the input voltage; the meter, however, has been calibrated in r.m.s. values. The calibration is correct only for sinusoidal voltages. Although the output voltage of the amplifier A_0 shows only little distortion, a simple low-pass filter F_0 has been placed behind this amplifier; in this way the distortion of the voltage across the meter GM 6016 is kept lower than 0.1%.

Checking the frequency characteristic

Via a coupling coil a voltage of 300 mV is obtained from each of the twelve oscillators of panel 5. These coupling coils are connected to two selector switches (S_1 and S_2, *fig. 10*) on panel 4 by coaxial cables; with these switches a measuring

provided with a sliding core of Ferroxcube; see the article referred to in note [7]), page 209).

Each of the twelve oscillators is provided with a potentiometer ($P_1 ... P_{12}$, fig. 9) with which the reference voltage and consequently the output voltages can be varied.

The way in which the equality of the measuring voltages is checked, requires further comment. A direct voltage is derived from the alternating voltage by means of a germanium diode Ge_1. A vibrating contact (S_4, fig. 10) converts this into a low-frequency pulsating voltage which is amplified by the amplifier A_1 and measured in a relative measure by the meter M. At the lower frequencies (1 to 1000 kc/s), from which a choice can be made by means of the selector switch S_1, a cathode follower is incorporated behind S_1. The germanium diode is connected direct to this cathode follower, and the meter GM 6016, which has to be checked, via a coaxial cable. However, at frequencies exceeding 1 Mc/s the difference between the voltages at the beginning and the end of the cable would become excessive (see the article mentioned in note [7]), page 214); therefore, at these higher frequencies, use is made of a germanium diode Ge_2 mounted in a probe at the same end of the cable where the meter is connected.

The frequency 1000 kc/s = 1 Mc/s occurs both in the lower and in the higher frequency range, so as to ensure a good agreement between the measurements in the two ranges.

By means of an attenuator a voltage of 3 mV is derived from the voltage of 300 mV which is induced in each of the coupling coils, for checking the meter without using the attenuator Z_2 belonging to it. For the frequencies from 1 to 1000 kc/s the first-mentioned attenuator consists of resistors (r_1, r_2, fig. 10), which are contained in the measuring rack; for the frequencies 1 to 30 Mc/s it consists of capacitors (C_1, C_2), which are mounted in the probe at the end of cable K_2.
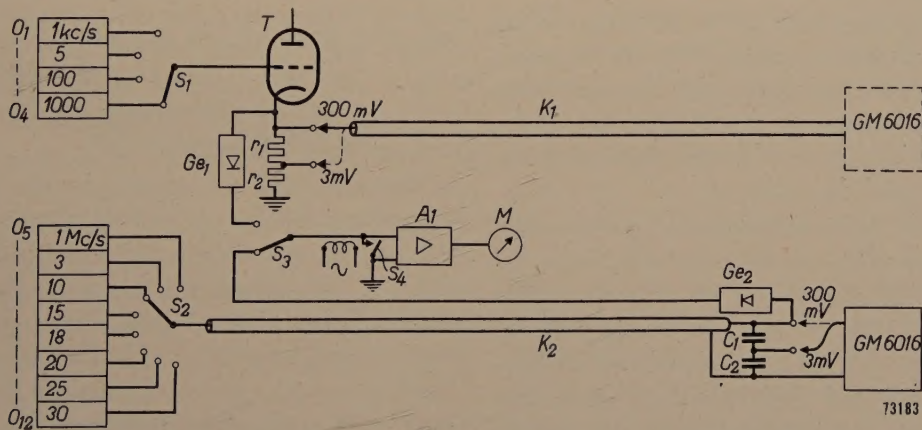

Fig. 10. Set-up for checking the freqeuncy characteristic of the millivoltmeter GM 6016. O_1-O_{12} oscillators with automatically stabilised voltages, with frequencies 1 kc/s to 30 Mc/s. S_1 selector switch 1-1000 kc/s. S_2 selector switch 1-30 Mc/s. S_3 selector switch (1-1000 kc/s)/ (1-30 Mc/s). T cathode follower. r_1, r_2 and C_1, C_2 voltage dividers, by which not only 300 mV but also 3 mV are available. K_1, K_2, coaxial cables. Ge_1, Ge_2 rectifying circuits with germanium diode (in the form of a probe). S_4 vibrating contact. A_1 amplifier. M meter.

voltage in either frequency range 1-1000 kc/s or 1-30 Mc/s is selected. With equal measuring voltages, the meter GM 6016 should give the same deflection within certain tolerances; should it fail to do so, then certain components of the meter should be readjusted until the frequency characteristic has become sufficiently flat (to facilitate this, certain coils are

Summary: In accordance with a well-known principle of regulating technique, the output voltage of a valve oscillator may be stabilised by the mutual comparison of two currents, one of which is obtained by rectifying the output voltage, whereas the other originates from a reference voltage source; from the difference between these currents a regulating voltage is derived, which serves for biasing the oscillator valve.

In this way, the magnitude of the output voltage can, with the aid of only a few extra circuit elements, be made highly insensitive to mains voltage fluctuations and to changes in the mutual conductance of the valve and of the impedance of the oscillatory circuit. The output voltage is less distorted than with an ordinary LC oscillator.

By making the reference voltage controllable, it becomes possible to adjust the output voltage to the value required, and by superimposing an alternating voltage on the reference voltage, the amplitude of the output voltage can be modulated. The upper limit of the modulation frequency and that of the modulation depth are discussed.

The tendency towards squegging can be counteracted by giving the correct value to certain capacitances. For the oscillator to start smoothly after switching on, care should be taken to prevent the current supplied by the reference voltage source from flowing as control grid current through the valve; to this end an auxiliary diode can be inserted in the circuit.

The valve EBF 80 contains a pentode system, which may serve as an oscillator valve, and two diode systems, one of which can rectify the alternating voltage to be stabilized, while the other may function as the auxiliary diode just mentioned.

A calibrating system for the type GM 6016 electronic millivoltmeter (1000 c/s to 30 Mc/s) is discussed as a practical application.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE
N.V. PHILIPS' GLOEILAMPENFABRIEKEN

Reprints of these papers not marked with an asterisk * can be obtained free of charge upon application to the ᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷᚷ *Administration of the Research Laboratory, Kastanjelaan, Eindhoven, Netherlands.*

2012: R. van der Veen: Fluorescence and induction phenomena in photosynthesis (Physiologia plantarum **4**, 486-494, 1951).

Fluorescence of chlorophyll in leaves shows induction phenomena during the first few minutes of illumination, which are very much like those of photosynthesis. The influence of preceding dark period, of temperature, of light intensity and of CO_2 and O_2 pressure on the adaptation curve was studied. An increase of photosynthesis during adaptation is nearly always correlated with a decrease of fluorescence.

The only exceptions are: (a) in the temperature limited range of photosynthesis, a lower temperature causes less photosynthesis, but does not result in any marked change in fluorescence and b) a high O_2 pressure causes both a lower rate of photosynthesis and lower fluorescence.

The results of these experiments are therefore mostly in agreement with the view of Katz; the fluorescence of chlorophyll can often be used as a flow meter for the energy of photosynthesis. Sometimes, however, photosynthesis and fluorescence do not show any correlation. In these cases it is probable that the light energy is used for the splitting of H_2O but that afterwards back reactions cause a reduction of the photosynthesis.

2013: W. de Groot: On the definition of standard illuminant A (Physica **17**, 920-922, 1951, No. 10).

In the Stockholm meeting 1951, the C. I. E. provisionally defined standard illuminant A by the value $\beta = c_2/T = 5.0386$ ($= 14350/2848 \approx 14380/2854$) to be used in Planck's (or Wien's) radiation formula. It is pointed out that if one wishes to realise the standard (in Wien's approximation) by constructing a black body, the monochromatic power ratio of which (with respect to a black body at the melting point of gold) is $\exp(a/\lambda)$, where $a = c_2/T_{Au} - \beta$, a change in c_2 and T_{Au} in general influences a, if β is kept constant. The spectral distribution of this source will in general deviate from the calculated value unless the "true" values of c_2 and T_{Au} are used in calculating a.

2014: A. Bierman: A systematic method with new standards to determine the correct magnitude of radiographical exposures (Acta radiologica **36**, 311-323, Oct. 1951).

The value of a radiographical exposure E is determined by $E = (kV/100)^5 mA \cdot sec/m^2$ (in what the writer calls "radiographic exposure units" or "r e u"). A standard object (a water phantom with definite dimensions) is used to characterise the radiographical outfit with the "specific exposure" E_s. This is the value of E required to obtain a film density of unity when radiographing the standard object. The exposure required for a certain object can be computed with the aid of E_s and "relative exposure" E/E_s. The values of of E/E_s appropriate to the various objects constitute a "basic exposure table", from which the correct exposure is derived by means of the formula $E = (E/E_s) \times E_s$. All computations can be avoided by the use of two nomograms.